

A. Y. KHINCHIN

**Mathematical
Foundations
of
INFORMATION
THEORY**

information theory



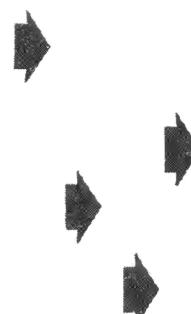


mathematical foundations of information theory

A. I. Khinchin

translated by
R. A. Silverman
institute of mathematical sciences
new york university
and
M. D. Friedman
lincoln laboratory
massachusetts institute of technology

dover publications, inc., new york



Copyright © 1957 by Dover Publications, Inc.

This new Dover edition is a translation of the
following two papers by A. I. KHINCHIN :

The Entropy Concept in Probability Theory

On the Fundamental Theorems of Information Theory.

Library of Congress Catalog Card Number: 57-13025

Manufactured in the United States of America.

CONTENTS

The Entropy Concept in Probability Theory

# 1. Entropy of Finite Schemes	2
# 2. The Uniqueness Theorem	9
# 3. Entropy of Markov chains	13
# 4. Fundamental Theorems	16
# 5. Application to Coding Theory	23

On the Fundamental Theorems of Information Theory

INTRODUCTION	30
CHAPTER I. Elementary Inequalities	34
# 1. Two generalizations of Shannon's inequality	34
# 2. Three inequalities of Feinstein	39
CHAPTER II. Ergodic Sources	44
# 3. Concept of a source. Stationarity. Entropy.....	44
# 4. Ergodic Sources	49
# 5. The E property. McMillan's theorem.....	54
# 6. The martingale concept. Doob's theorem.....	58
# 7. Auxiliary propositions	64
# 8. Proof of McMillan's theorem.....	70
CHAPTER III. Channels and the sources driving them	75
# 9. Concept of channel. Noise. Stationarity. Anticipation and memory.....	75
# 10. Connection of the channel to the source	78
# 11. The ergodic case	85
CHAPTER IV. Feinstein's Fundamental Lemma.....	90
# 12. Formulation of the problem	90
# 13. Proof of the lemma	93
CHAPTER V. Shannon's Theorems	102
# 14. Coding	102
# 15. The first Shannon theorem	104
# 16. The second Shannon theorem	109
CONCLUSION	111
REFERENCES	120

The Entropy Concept in Probability Theory

(Uspekhi Matematicheskikh Nauk, vol. VIII, no. 3, 1953, pp. 3-20)

. In his article “On the Drawing of Maps” P. L. Chebyshev beautifully expresses the nature of the relation between scientific theory and practice (discussing the case of mathematics): “The bringing together of theory and practice leads to the most favorable results; not only does practice benefit, but the sciences themselves develop under the influence of practice, which reveals new subjects for investigation and new aspects of familiar subjects.” A striking example of the phenomenon described by Chebyshev is afforded by the concept of entropy in probability theory, a concept which has evolved in recent years from the needs of practice. This concept first arose in attempting to create a theoretical model for the transmission of information of various kinds. In the beginning the concept was introduced in intimate association with transmission apparatus of one kind or another; its general theoretical significance and properties, and the general nature of its application to practice were only gradually realized. As of the present, a unified exposition of the theory of entropy can be found only in specialized articles and monographs dealing with the transmission of information. Although the study of entropy has actually evolved into an important and interesting chapter of the general theory of probability, a presentation of it in this general theoretical setting has so far been lacking.

This article represents a first attempt at such a presentation. In writing it, I relied mainly on Shannon's paper "The Mathematical Theory of Communication".* However, Shannon's treatment is not always sufficiently complete and mathematically correct, so that besides having to free the theory from practical details, in many instances I have amplified and changed both the statement of definitions and the statement and proofs of theorems. There is no doubt that in the years to come the study of entropy will become a permanent part of probability theory; the work I have done seems to me to be a necessary stage in the development of this study.

#1. Entropy of Finite Schemes

In probability theory a *complete system of events* A_1, A_2, \dots, A_n means a set of events such that one and only one of them must occur at each trial (e.g., the appearance of 1, 2, 3, 4, 5, or 6 points in throwing a die). In the case $n=2$ we have a simple alternative or pair of *mutually exclusive* events (e.g., the appearance of heads or tails in tossing a coin). If we are given the events A_1, A_2, \dots, A_n of a complete system, together with their probabilities p_1, p_2, \dots, p_n ($p_i \geq 0, \sum_{i=1}^n p_i = 1$), then we say that we have a *finite scheme*

$$A = \begin{pmatrix} A_1 & A_2 & \cdots & A_n \\ p_1 & p_2 & \cdots & p_n \end{pmatrix}. \quad (1)$$

In the case of a "true" die, designating the appearance of i points by A_i ($1 \leq i \leq 6$), we have the finite scheme

$$\begin{pmatrix} A_1 & A_2 & A_3 & A_4 & A_5 & A_6 \\ 1/6 & 1/6 & 1/6 & 1/6 & 1/6 & 1/6 \end{pmatrix}.$$

* C. E. Shannon, Bell System Technical Journal, 27, 379-423; 623-656 (1948).

Every finite scheme describes a state of *uncertainty*. We have an experiment, the outcome of which must be one of the events A_1, A_2, \dots, A_n , and we know only the probabilities of these possible outcomes. It seems obvious that the amount of uncertainty is different in different schemes. Thus, in the two simple alternatives

$$\begin{pmatrix} A_1 & A_2 \\ 0.5 & 0.5 \end{pmatrix}, \quad \begin{pmatrix} A_1 & A_2 \\ 0.99 & 0.01 \end{pmatrix},$$

the first obviously represents much more uncertainty than the second; in the second case, the result of the experiment is "almost surely" A_1 , while in the first case we naturally refrain from making any predictions. The scheme

$$\begin{pmatrix} A_1 & A_2 \\ 0.3 & 0.7 \end{pmatrix}$$

represents an amount of uncertainty intermediate between the preceding two, etc.

For many applications it seems desirable to introduce a quantity which in a reasonable way measures the amount of uncertainty associated with a given finite scheme. We shall see that the quantity

$$H(p_1, p_2, \dots, p_n) = - \sum_{k=1}^n p_k \lg p_k,$$

can serve as a very suitable measure of the uncertainty of the finite scheme (1); the logarithms are taken to an arbitrary but fixed base, and we always take $p_k \lg p_k = 0$ if $p_k = 0$. We shall call the quantity $H(p_1, p_2, \dots, p_n)$ the *entropy* of the finite scheme (1), pursuing a physical analogy which there is no need to go into here. We now convince ourselves that this function actually has a number of properties which we might expect

of a reasonable measure of uncertainty of a finite scheme.

First of all, we see immediately that $H(p_1, p_2, \dots, p_n) = 0$, if and only if one of the numbers p_1, p_2, \dots, p_n is one and all the others are zero. But this is just the case where the result of the experiment can be predicted beforehand with complete certainty, so that there is no uncertainty as to its outcome. In all other cases the entropy is positive.

Furthermore, for fixed n it is obvious that the scheme with the most uncertainty is the one with equally likely outcomes, i.e., $p_k = 1/n$ ($k = 1, 2, \dots, n$), and in fact the entropy assumes its largest value for just these values of the variables p_k . The easiest way to see this is to use an inequality which is valid for any continuous convex function $\varphi(x)$

$$\varphi\left(\frac{1}{n} \sum_{k=1}^n a_k\right) \leq \frac{1}{n} \sum_{k=1}^n \varphi(a_k),$$

where a_1, a_2, \dots, a_n are any positive numbers. Setting $a_k = p_k$ and $\varphi(x) = x \lg x$, and bearing in mind that $\sum_{k=1}^n p_k = 1$, we find

$$\varphi\left(\frac{1}{n}\right) = \frac{1}{n} \lg \frac{1}{n} \leq \frac{1}{n} \sum_{k=1}^n p_k \lg p_k = -\frac{1}{n} H(p_1, p_2, \dots, p_n),$$

whence

$$H(p_1, p_2, \dots, p_n) \leq \lg n = H\left(\frac{1}{n}, \frac{1}{n}, \dots, \frac{1}{n}\right), \quad \text{Q.E.D.}$$

Suppose now we have two finite schemes

$$A = \begin{pmatrix} A_1 & A_2 & \cdots & A_n \end{pmatrix}, \quad B = \begin{pmatrix} B_1 & B_2 & \cdots & B_m \end{pmatrix},$$

and let these two schemes be (mutually) independent, i.e., the probability π_{kl} of the joint occurrence of the events A_k and B_l is $p_k q_l$. Then, the set of events $A_k B_l$ ($1 \leq k \leq n, 1 \leq l \leq m$),

with probabilities π_{kl} represents another finite scheme, which we call the *product* of the schemes A and B and designate by AB . Let $H(A)$, $H(B)$, and $H(AB)$ be the corresponding entropies of the schemes A , B , and AB . Then

$$H(AB) = H(A) + H(B), \quad (2)$$

for, in fact

$$\begin{aligned} -H(AB) &= \sum_k \sum_l \pi_{kl} \lg \pi_{kl} = \sum_k \sum_l p_k q_l (\lg p_k + \lg q_l) = \\ &= \sum_k p_k \lg p_k \sum_l q_l + \sum_l q_l \lg q_l \sum_k p_k = -H(A) - H(B). \end{aligned}$$

We now turn to the case where the schemes A and B are (mutually) dependent. We denote by q_{kl} the probability that the event B_l of the scheme B occurs, given that the event A_k of the scheme A occurred, so that

$$\pi_{kl} = p_k q_{kl} \quad (1 \leq k \leq n, 1 \leq l \leq m).$$

Then

$$\begin{aligned} -H(AB) &= \sum_k \sum_l p_k q_{kl} (\lg p_k + \lg q_{kl}) = \\ &= \sum_k p_k \lg p_k \sum_l q_{kl} + \sum_k p_k \sum_l q_{kl} \lg q_{kl}. \end{aligned}$$

Here $\sum_l q_{kl} = 1$ for any k , and the sum $-\sum_l q_{kl} \lg q_{kl}$ can be regarded as the conditional entropy $H_k(B)$ of the scheme B , calculated on the assumption that the event A_k of the scheme A occurred. We obtain

$$H(AB) = H(A) + \sum_k p_k H_k(B).$$

The conditional entropy $H_k(B)$ is obviously a random variable in the scheme A ; its value is completely determined by the knowledge of which event A_k of the scheme A actually occurred. Therefore, the last term of the right side is the *mathematical*

expectation of the quantity $H(B)$ in the scheme A , which we shall designate by $H_A(B)$. Thus in the most general case, we have

$$H(AB) = H(A) + H_A(B). \quad (3)$$

It is self-evident that the relation (3) reduces to (2) in the special case where the schemes A and B are independent.

It is also interesting to note that in all cases $H_A(B) \leq H(B)$. It is reasonable to interpret this inequality as saying that, on the average, knowledge of the outcome of the scheme A can only decrease the uncertainty of the scheme B . To prove this, we observe that any continuous convex function $f(x)$ obeys the inequality*

$$\sum_k \lambda_k f(x_k) \geq f\left(\sum_k \lambda_k x_k\right),$$

if $\lambda_k \geq 0$ and $\sum_k \lambda_k = 1$. Therefore, setting $f(x) = x \lg x$, $\lambda_k = p_k$, $x_k = q_{kl}$, we find for arbitrary l that

$$\sum_k p_k q_{kl} \lg q_{kl} \geq \left(\sum_k p_k q_{kl}\right) \lg \left(\sum_k p_k q_{kl}\right) = q_l \lg q_l,$$

since obviously $\sum_k p_k q_{kl} = q_l$. Summing over l , we obtain on the left side the quantity

$$\sum_k p_k \sum_l q_{kl} \lg q_{kl} = -\sum_k p_k H_k(B) = -H_A(B),$$

and consequently we find

$$-H_A(B) \geq \sum_l q_l \lg q_l = -H(B), \quad \text{Q.E.D.}$$

If we carry out an experiment the possible outcomes of which are described by the given scheme A , then in doing so we obtain some *information* (i.e., we find out which of the events

* See, for example, Hardy, Littlewood, and Polya, *Inequalities*, Cambridge University Press, 1934.

A_k actually occurs), and the uncertainty of the scheme is completely eliminated. Thus, we can say that the information given us by carrying out some experiment consists in removing the uncertainty which existed before the experiment. The larger this uncertainty, the larger we consider to be the amount of information obtained by removing it. Since we agreed to measure the uncertainty of a finite scheme A by its entropy $H(A)$, it is natural to express the amount of information given by removing this uncertainty by an increasing function of the quantity $H(A)$. The choice of this function means the choice of some unit for the quantity of information and is therefore fundamentally a matter of indifference. However, the properties of entropy which we demonstrated above show that it is especially convenient to take this quantity of information proportional to the entropy. Indeed, consider two finite schemes A and B and their product AB . Realization of the scheme AB is obviously equivalent to realization of both of the schemes A and B . Therefore, if the two schemes A and B are independent, it is natural to require the information given by the realization of the scheme AB to be the sum of the two amounts of information given by the realization of the schemes A and B ; since in this case

$$H(AB) = H(A) + H(B),$$

this requirement will actually be met, if we consider the amount of information given by the realization of a finite scheme to be proportional to the entropy of the scheme. Of course, the constant of proportionality can be taken as unity, since this choice corresponds merely to a choice of units. Thus, in all that follows, we can consider the amount of information given

by the realization of a finite scheme to be equal to the entropy of the scheme. This stipulation makes the concept of entropy especially significant for information theory.

In view of this stipulation, let us consider the case of two dependent schemes A and B and the corresponding relation (3). The amount of information given by the realization of the scheme AB is equal to $H(AB)$. However, as explained above, in the general case, this cannot be equal to $H(A)+H(B)$. Indeed, consider the extreme case where knowledge of the outcome of the scheme A also determines with certainty the outcome of the scheme B , so that each event A_k of the scheme A can occur only in conjunction with a specific event B_l of the scheme B . Then, after realization of the scheme A , the scheme B completely loses its uncertainty, and we have $H_A(B)=0$; moreover, in this case realization of the scheme B obviously gives no further information, and we have $H(AB)=H(A)$, so that relation (3) is indeed satisfied. In all cases, the quantity $H_k(B)$ introduced above is the amount of information given by the scheme B , given that the event A_k occurred in the scheme A ; therefore the quantity $H_A(B)=\sum_k p_k H_k(B)$ is the mathematical expectation of the amount of additional information given by realization of the scheme B after realization of scheme A and reception of the corresponding information. Therefore, the relation (3) has the following very reasonable interpretation: *The amount of information given by the realization of the two finite schemes A and B , equals the amount of information given by the realization of scheme A , plus the mathematical expectation of the amount of additional information given by the realization of scheme B after the realization of the scheme A .* In just the same way we can give an

entirely reasonable interpretation of the general inequality $H_A(B) \leq H(B)$ proved above: *The amount of information given by the realization of a scheme B can only decrease if another scheme A is realized beforehand.*

#2. The Uniqueness Theorem

Among the properties of entropy which we have proved, we can consider the following two as basic:

1. For given n and for $\sum_{k=1}^n p_k = 1$, the function $H(p_1, p_2, \dots, p_n)$ takes its largest value for $p_k = \frac{1}{n}$ ($k=1, 2, \dots, n$).
2. $H(AB) = H(A) + H_A(B)$.

We add to these two properties a third, which obviously must be satisfied by any reasonable definition of entropy. Since the schemes

$$\begin{pmatrix} A_1 & A_2 & \cdots & A_n \\ p_1 & p_2 & \cdots & p_n \end{pmatrix} \text{ and } \begin{pmatrix} A_1 & A_2 & \cdots & A_n & A_{n+1} \\ p_1 & p_2 & \cdots & p_n & 0 \end{pmatrix},$$

are obviously not substantively different, we must have

3. $H(p_1, p_2, \dots, p_n, 0) = H(p_1, p_2, \dots, p_n)$. (Adding the impossible event or any number of impossible events to a scheme does not change its entropy.)

We now prove the following important proposition:

Theorem 1.

Let $H(p_1, p_2, \dots, p_n)$ be a function defined for any integer n and for all values p_1, p_2, \dots, p_n such that $p_k \geq 0$ ($k=1, 2, \dots, n$), $\sum_{k=1}^n p_k = 1$. If for any n this function is continuous with respect to all its arguments, and if it has the properties 1, 2, and 3, then

$$H(p_1, p_2, \dots, p_n) = -\lambda \sum_{k=1}^n p_k \lg p_k,$$

where λ is a positive constant.

This theorem shows that the expression for the entropy of a finite scheme which we have chosen is the only one possible if we want it to have certain general properties which seem necessary in view of the actual meaning of the concept of entropy (as a measure of uncertainty or as an amount of information).

Proof.

For brevity we set

$$H\left(\frac{1}{n}, \frac{1}{n}, \dots, \frac{1}{n}\right) = L(n);$$

we shall show that $L(n) = \lambda \lg n$, where λ is a positive constant. By 3 and 1, we have

$$L(n) = H\left(\frac{1}{n}, \frac{1}{n}, \dots, \frac{1}{n}, 0\right) \leq H\left(\frac{1}{n+1}, \frac{1}{n+1}, \dots, \frac{1}{n+1}\right) = L(n+1),$$

so that $L(n)$ is a non-decreasing function of n . Let m and r be positive integers. Consider m mutually independent finite schemes S_1, S_2, \dots, S_m , each of which contains r equally likely events, so that

$$H(S_k) = H\left(\frac{1}{r}, \frac{1}{r}, \dots, \frac{1}{r}\right) = L(r) \quad (1 \leq k \leq m).$$

By Property 2 (generalized to the case of m schemes) we have, in view of the independence of the schemes S_k

$$H(S_1 S_2 \cdots S_m) = \sum_{k=1}^m H(S_k) = m L(r).$$

But the product scheme $S_1 S_2 \cdots S_m$ obviously consists of r^m equally likely events, so that its entropy is $L(r^m)$. Therefore we have

$$L(r^m) = m L(r), \quad (4)$$

and similarly, for any other pair of positive integers n and s

$$L(s^n) = n L(s). \quad (5)$$

Now let the numbers r , s , and n be given arbitrarily, but let the number m be determined by the inequalities

$$r^m \leq s^n \leq r^{m+1}, \quad (6)$$

whence

$$\begin{aligned} m \lg r &\leq n \lg s < (m+1) \lg r, \\ \frac{m}{n} &\leq \frac{\lg s}{\lg r} < \frac{m}{n} + \frac{1}{n} \end{aligned} \quad (7)$$

It follows from (6) by the monotonicity of the function $L(n)$ that

$$L(r^m) \leq L(s^n) \leq L(r^{m+1}),$$

and, consequently, by (4) and (5)

$$m L(r) \leq n L(s) \leq (m+1) L(r),$$

so that

$$\frac{m}{n} \leq \frac{L(s)}{L(r)} \leq \frac{m}{n} + \frac{1}{n}. \quad (8)$$

Finally, it follows from (7) and (8) that

$$\left| \frac{L(s)}{L(r)} - \frac{\lg s}{\lg r} \right| \leq \frac{1}{n}$$

Since the left side of this inequality is independent of m , and since n can be chosen arbitrarily large in the right side

$$\frac{L(s)}{\lg s} = \frac{L(r)}{\lg r},$$

which, in view of the arbitrariness of r and s , means that

$$L(n) = \lambda \lg n,$$

where λ is a constant. By the monotonicity of the function $L(n)$, we have $\lambda \geq 0$, and our assertion is proved.

This assertion represents the special case $p_k = 1/n$ ($1 \leq k \leq n$) of the theorem to be proved. We now consider the more general case, where the p_k ($k = 1, 2, \dots, n$) are any rational numbers. Let

$$p_k = \frac{g_k}{g} \quad (k = 1, 2, \dots, n),$$

where all the g_k are positive integers and $\sum_{k=1}^n g_k = g$. Let the finite scheme A consist of n events with probabilities p_1, p_2, \dots, p_n . Our problem consists in defining the entropy of this scheme. To this end, we consider a second scheme B , which is dependent on A and is defined as follows: The scheme B contains g events B_1, B_2, \dots, B_g , which we devide into n groups, containing g_1, g_2, \dots, g_n events, respectively. If the event A_k occurred in scheme A , then in scheme B all the g_k events of the k 'th group have the same probability $1/g_k$, and all the events of the other groups have probability zero (are impossible). Thus, given any outcome A_k of the scheme A , the scheme B reduces to a system of g_k equally likely events, so that the conditional entropy

$$H_k(B) = H(1/g_k, 1/g_k, \dots, 1/g_k) = L(g_k) = \lambda \lg g_k,$$

which means that

$$H_A(B) = \sum_{k=1}^n p_k H_k(B) = \lambda \sum_{k=1}^n p_k \lg g_k = \lambda \sum_{k=1}^n p_k \lg p_k + \lambda \lg g. \quad (9)$$

We return now to the product scheme AB , consisting of the

events $A_k B_l$ ($1 \leq k \leq n$, $1 \leq l \leq g$). Such an event is possible only if B_l belongs to the k 'th group. Thus, the number of possible events $A_k B_l$ for a given k is g_k , and the total number of possible events in the scheme AB is $\sum_{k=1}^n g_k = g$. The probability of each possible event $A_k B_l$ is obviously $p_k/g_k = 1/g$, i.e., is the same for all the events. Thus, the scheme AB consists of g equally likely events, and therefore

$$H(AB) = L(g) = \lambda \lg g.$$

Using property (2) and relation (9), we find

$$\lambda \lg g = H(A) + \lambda \sum_{k=1}^n p_k \lg p_k + \lambda \lg g,$$

whence

$$H(A) = H(p_1, p_2, \dots, p_n) = -\lambda \sum_{k=1}^n p_k \lg p_k. \quad (10)$$

Finally, relation (10) which we have proved for rational p_1, p_2, \dots, p_n , must be valid for any values of its arguments because of the postulated continuity of the function $H(p_1, p_2, \dots, p_n)$. Thus the proof of Theorem 1 is complete.

#3. Entropy of Markov chains

Suppose we have a simple stationary Markov chain with a finite number of states A_1, A_2, \dots, A_n and with the transition probability matrix p_{ik} ($i, k = 1, 2, \dots, n$). We denote by P_k the probability of the state A_k ($1 \leq k \leq n$), so that in particular

$$\sum_{k=1}^n P_k p_{kl} = P_l \quad (l = 1, 2, \dots, n). \quad (11)$$

If the system is in state A_i , then its transitions to the different states A_k ($k = 1, 2, \dots, n$) form a finite scheme

$$\begin{pmatrix} A_1 & A_2 & \cdots & A_n \\ p_{i1} & p_{i2} & \cdots & p_{in} \end{pmatrix},$$

the entropy of which

$$H_i = - \sum_{k=1}^n p_{ik} \lg p_{ik}$$

depends on i and can be regarded as a measure of the amount of information obtained when the Markov chain moves one step ahead, starting from the initial state A_i . The average of this quantity over all initial states, i.e., the quantity

$$H = \sum_{i=1}^n P_i H_i = - \sum_{i=1}^n \sum_{k=1}^n P_i p_{ik} \lg p_{ik},$$

is therefore to be regarded as a measure of the average amount of information obtained when the given Markov chain moves one step ahead. This quantity H , which we shall call the entropy of the chain in question obviously characterizes the chain as a whole; it is clear that it is uniquely determined by giving the state probabilities P_i and the transition probabilities p_{ik} ($1 \leq i \leq n$, $1 \leq k \leq n$).

All the concepts which are defined for moving one step ahead can be easily and naturally generalized to the case of moving ahead an arbitrary number of steps r . If the system is in state A_i , then it is easy to calculate the probability that in the next r trials we shall find it in the states $A_{k_1}, A_{k_2}, \dots, A_{k_r}$ in turn, where k_1, k_2, \dots, k_r are arbitrary numbers from 1 to n . Thus, the subsequent fate of a system initially in the state A_i in the next r trials is described by a finite scheme (with n^r events), with a definite entropy which we designate by $H_i^{(r)}$ and regard as a measure of the amount of information obtained in moving ahead r steps in the chain, starting from the initial state A_i . The quantity

$$H^{(r)} = \sum_{i=1}^n P_i H_i^{(r)}$$

is to be regarded as the *average amount of information given by moving ahead r steps in the given Markov chain*. We shall call it the r -step entropy of the chain in question. The one-step entropy defined above can obviously be written as $H^{(1)}$.

If the notion of the quantity $H^{(r)}$ as the average amount of information obtained in moving ahead r steps in a given Markov chain is to be reasonable, then it is natural to require that for arbitrary positive integers r and s we have

$$H^{(r+s)} = H^{(r)} + H^{(s)},$$

or, equivalently, $H^{(r)} = rH^{(1)}$ for any positive integer r . It is easily seen that this is actually the case. In the first place, the relation $H^{(r)} = rH^{(1)}$ is trivially true for $r=1$. Suppose now that it is true for some $r \geq 1$; we shall show that in this case $H^{(r+1)} = (r+1)H^{(1)}$.

Let the system be in the state A_i ; the finite scheme which describes the fate of the system in the next $r+1$ trials, can then be regarded as the product of two dependent schemes:

A) the scheme corresponding to the immediately following trial with the entropy $H_i^{(1)}$ and

B) the scheme describing the fate of the system in the next r trials; the entropy of this scheme is $H_k^{(r)}$, if the outcome of scheme A was the event A_k . According to the general relation

$$H(AB) = H(A) + H_A(B),$$

we have

$$H_i^{(r+1)} = H_i^{(1)} + \sum_{k=1}^n p_{ik} H_k^{(r)},$$

and consequently, in view of (11)

$$\begin{aligned}
 H^{(r+1)} &= \sum_{i=1}^n P_i H_i^{(r+1)} = \sum_{i=1}^n P_i H_i^{(1)} + \sum_{k=1}^n H_k^{(1)} \sum_{i=1}^n P_i p_{ik} = \\
 &= H^{(1)} + \sum_{k=1}^n P_k H_k^{(1)} = H^{(1)} + H^{(r)} = (r+1)H^{(1)}, \quad \text{Q.E.D.}
 \end{aligned}$$

In a large number of cases, the application of the concept of the entropy of Markov chains in information theory is based on two fundamental theorems, which we shall prove in the next section.

#4. Fundamental Theorems

Suppose now that the Markov chain we are studying obeys the law of large numbers, i.e. that in a sufficiently long sequence of s consecutive trials the relative frequency $\frac{m_i}{s}$ of occurrence of the state A_i will differ from P_i by an arbitrarily small amount, with a probability arbitrarily close to unity. In other words, for arbitrarily small $\varepsilon > 0$ and $\delta > 0$, and for sufficiently large s

$$P\left\{\left|\frac{m_i}{s} - P_i\right| > \delta\right\} < \varepsilon^*). \quad (12)$$

For brevity, we shall call such a chain *ergodic*.

Each possible result of the series of s consecutive trials of the given Markov chain can be written as a "sequence"

$$A_{k_1}, A_{k_2}, \dots, A_{k_s}, \quad (C)$$

where k_1, k_2, \dots, k_s are numbers from 1 to n . The probability of realizing the sequence (C) does not depend on the part of the chain where the series of trials begins (because of the

*) A. A. Markov showed that for this to be the case it is sufficient to assume that the chain in question is "transitive", i.e., that a transition is possible from any state to any other state in a sufficiently large number of steps. For the chains with which information theory is concerned, this hypothesis is apparently always fulfilled.

stationarity), and is obviously equal to

$$p(C) = P_{k_1} p_{k_1 k_2} p_{k_2 k_3} \cdots p_{k_{s-1} k_s}.$$

Let i and l be two arbitrary numbers from 1 to n , and let m_{il} be the number of pairs of the form $k_r k_{r+1}$ ($1 \leq r \leq s$) in which $k_r = i$, $k_{r+1} = l$. Then, clearly the probability of the sequence (C) can be written in the form

$$p(C) = P_{k_1} \prod_{i=1}^n \prod_{l=1}^n (p_{il})^{m_{il}}. \quad (13)$$

Theorem 2.

Given $\varepsilon > 0$ and $\eta > 0$, no matter how small, for sufficiently large n all sequences of the form (C) can be divided into two groups with the following properties: 1) the probability $p(C)$ of any sequence of the first group satisfies the inequality

$$\left| \frac{\lg \frac{1}{p(C)}}{s} - H \right| < \eta, \quad (14)$$

and 2) the sum of the probabilities of all sequences of the second group is less than ε .

In other words, all sequences with the exception of a very low probability group have probabilities lying between $a^{-s(H+\eta)}$ and $a^{-s(H-\eta)}$, where a is the base of the system of logarithms used. Here H is the one-step entropy of the given chain.

Proof.

We shall agree to assign the sequence (C) to the first group if it has the following two properties:

1. it is a possible outcome, i.e. $p(C) > 0$, and
2. for any i, l ($1 \leq i \leq n, 1 \leq l \leq n$) the inequality

$$|m_{il} - sP_i p_{il}| < s\delta \quad (15)$$

obtains. All the other sequences are assigned to the second group. We shall show that this division satisfies both requirements of theorem 2, if $\delta > 0$ is sufficiently small and if s is sufficiently large.

1) Suppose the sequence (C) belongs to the first group. It follows from (15) that

$$m_{il} = sP_i p_{il} + s\delta\theta_{il}, \quad |\theta_{il}| < 1, \quad 1 \leq i \leq n, \quad 1 \leq l \leq n.$$

We now substitute these expressions for the numbers m_{il} in (13), where in doing so we must bear in mind that the requirement that the given sequence be a possible outcome implies that $m_{il} = 0$ when $p_{il} = 0$. Thus in the product (13) we must restrict ourselves to factors such that $p_{il} > 0$, which we shall denote by an asterisk on the product signs. We find

$$\begin{aligned} p(C) &= P_{k_1} \prod_i \prod_l^* (p_{il})^{sP_i p_{il} + s\delta\theta_{il}}, \\ \lg \frac{1}{p(C)} &= -\lg P_{k_1} - s \sum_i \sum_l^* P_i p_{il} \lg p_{il} - s\delta \sum_i \sum_l^* \theta_{il} \lg p_{il} = \\ &= -\lg P_{k_1} + sH - s\delta \sum_i \sum_l^* \theta_{il} \lg p_{il}, \end{aligned}$$

whence

$$\left| \frac{\lg \frac{1}{p(C)}}{s} - H \right| < \frac{1}{s} \lg \frac{1}{P_{k_1}} + \delta \sum_i \sum_l^* \lg \frac{1}{p_{il}}.$$

This means that for sufficiently large s

$$\left| \frac{\lg \frac{1}{p(C)}}{s} - H \right| < \eta,$$

where $\eta > 0$ is as small as we please if δ is sufficiently small.

Thus, the first requirement of Theorem 2 is satisfied.

2) Turning now to the calculation of the sum of the probabilities of all the sequences of the second group, we note first of all that the impossible sequences in this group are not included in this calculation, since their probabilities are zero. Thus, we only have to calculate the sum of the probabilities of those sequences for which the inequality (15) does not hold for at least one pair of indices i, l ; to do this it suffices to calculate the quantity

$$\sum_{i=1}^n \sum_{l=1}^n P \{ |m_{il} - sP_i p_{il}| \geq s\delta \}.$$

First we fix the indices i and l . By (12) we have for sufficiently large s

$$P \left\{ |m_i - sP_i| < \frac{\delta}{2}s \right\} > 1 - \varepsilon.$$

If the inequality $\{ \}$ is satisfied, and if s is sufficiently large, then m_i is as large as we please, and therefore by Bernoulli's theorem

$$P \left\{ \left| \frac{m_{il}}{m_i} - p_{il} \right| < \frac{\delta}{2} \right\} > 1 - \varepsilon.$$

Therefore the probability of satisfying both of the inequalities

$$|m_i - sP_i| < \frac{\delta}{2}s, \quad (16)$$

$$|m_{il} - m_i p_{il}| < \frac{\delta}{2}m_i \leq \frac{\delta}{2}s \quad (17)$$

exceeds $(1 - \varepsilon)^2 > 1 - 2\varepsilon$. But it follows from (16) that

$$|p_{il}m_i - sP_i p_{il}| < p_{il} \frac{\delta}{2}s < \frac{\delta}{2}s,$$

and this together with (17) gives

$$|m_{il} - sP_i p_{il}| < \delta s. \quad (18)$$

Thus for any i and l , we have for sufficiently large s

$$P\{|m_{il} - sP_i p_{il}| < \delta s\} > 1 - 2\epsilon,$$

which means that

$$P\{|m_{il} - sP_i p_{il}| > \delta s\} < 2\epsilon.$$

From this it follows that

$$\sum_{i=1}^n \sum_{l=1}^n P\{|m_{il} - sP_i p_{il}| > \delta s\} < 2n^2\epsilon.$$

Since the right side of this inequality is together with ϵ arbitrarily small, the sum of the probabilities of all the sequences of the second group can be made as small as we please for sufficiently large s , i.e., the second requirement of Theorem 2 is also satisfied. Thus Theorem 2 is proved.

We now note that the number of all s -term sequences of the form (C) equals n^s , and we arrange these sequences in order of decreasing probability $p(C)$. We select sequences from this series in the order in which we have arranged them until the sum of the probabilities of the sequences selected just exceeds a preassigned positive number λ ($0 < \lambda < 1$). We denote by $N_s(\lambda)$ the number of sequences so selected. Theorem 2 permits us to make the following important estimate of the number $N_s(\lambda)$.

Theorem 3. $\lim_{s \rightarrow \infty} \frac{\lg N_s(\lambda)}{s} = H.$

In particular, the indicated limit does not depend on the number λ , if only $0 < \lambda < 1$ and λ remains constant as s increases.

Proof.

We agree to call a sequence (C) *standard* if its probability $p(C)$ satisfies the inequality (14) where η is a fixed, arbitrarily small positive number. By Theorem 2, the sum of the probabilities of all non-standard sequences is arbitrarily small for sufficiently large s . The inequality (14) is equivalent to the inequality

$$a^{-s(H+\eta)} < p(C) < a^{-s(H-\eta)}, \quad (19)$$

which therefore are characteristic of standard sequences (a denotes the base of the system of logarithms used).

The following sequences are among the selected sequences (with sum of probabilities $>\lambda$):

1) all non-standard sequences with probabilities $p(C) \geq a^{-s(H-\eta)}$; the sum of the probabilities of such sequences does not exceed the sum of the probabilities of all non-standard sequences, which latter is for sufficiently large s less than any $\varepsilon > 0$, however small.

2) a certain number $M_s(\lambda)$ of standard sequences, the sum of the probabilities of which must be greater than $\lambda - \varepsilon$ (since the sum of the probabilities of the selected sequences exceeds λ).

The non-standard sequences with probabilities $p(C) \leq a^{-s(H+\eta)}$ can not be among the selected sequences, since according to Theorem 2 the sum of the probabilities of the standard sequences by themselves exceeds λ . Thus, for all the selected sequences $p(C) > a^{-s(H+\eta)}$, and therefore the sum of the probabilities of the selected sequences is greater than $N_s(\lambda)a^{-s(H+\eta)}$. On the other hand, this sum is obviously less than $\lambda + q$, where q is the probability of the last sequence selected; thus

$$N_s(\lambda)a^{-s(H+\eta)} < \lambda + q < \lambda + a^{-s(H-\eta)},$$

which means (since $\lambda < 1$) that for sufficiently large s

$$N_s(\lambda) a^{-s(H+\eta)} < 1,$$

whence

$$\frac{\lg N_s(\lambda)}{s} < H + \eta. \quad (20)$$

On the other hand, the $M_s(\lambda)$ standard sequences which we have selected have probabilities $p(C) < a^{-s(H-\eta)}$, while their sum is larger than $\lambda - \varepsilon$. It follows that

$$M_s(\lambda) a^{-s(H-\eta)} > \lambda - \varepsilon,$$

so that *a fortiori*

$$N_s(\lambda) a^{-s(H-\eta)} > \lambda - \varepsilon,$$

whence

$$\frac{\lg N_s(\lambda)}{s} > H - \eta + \frac{1}{s} \lg (\lambda - \varepsilon). \quad (21)$$

Combining the inequalities (20) and (21) obviously proves Theorem 3, since the number η can be made arbitrarily small for sufficiently large s .

The great value of Theorem 3 for a variety of applications depends on the following considerations. Whereas the number of sequences of type (C) is $n^s = a^{s \lg n}$, the number $N_s(\lambda)$ of selected sequences is approximately a^{sH} , as Theorem 3 shows. If we recall that $\lg n$ is the maximum value of the entropy H and that, consequently, we always have $H < \lg n$ (except for a trivial case) and if we choose the number λ very close to unity, then Theorem 3 shows that a negligibly small fraction of all the sequences (C) has a sum of probabilities arbitrarily close to unity (for sufficiently large s). Moreover, we see that the *entropy* of the given Markov chain plays a decisive role in determining *how small* this fraction is.

#5 Application to Coding Theory

In order to give at least one example which illustrates the practical applications of the entropy concept, we now consider one of the simplest problems of coding theory. Suppose that the text which is to be coded consists of a sequence of symbols (letters) belonging to a finite set (alphabet), and denote by m the number of different symbols (so that the number of different sequences of length s is m^s). We shall regard this text as a simple Markov chain of the type considered in the two preceding sections and assume that its statistical structure is known, in particular its entropy H . (We know that the maximum value of H is $\lg m$.) We restrict ourselves to a consideration of the simplest case where the text at hand is coded *into the same alphabet*, i.e., each sequence from the text is coded into a sequence of letters from the same alphabet. (Actually, of course, it is only important that the coded text has the same *number* of symbols as the uncoded text, since no role is played by the designation of the symbols.) It goes without saying that the rules of coding must guarantee that the original text can be uniquely reconstructed from the coded text, which requires in particular that different sequences of the uncoded text must be coded differently.

It is immediately clear that by using as short a coding as possible for the most commonly encountered sequences, and conversely, by leaving the longer coding for the more rarely encountered sequences, we have the possibility of making the coded text shorter than the original, which obviously might constitute a practical and economic advantage. In order to analyze this possibility, we must first of all choose quantities

which can measure in a natural way this kind of compression by coding. It is apparent at once that both the possible amount of compression and the choice of an optimum code to achieve it depend entirely on the statistical structure of the given text.

Each s -term sequence (C) of symbols from the input text has a definite probability $p(C)$, and the sequence of coded text into which it is transformed by the coding has a definite length $\sigma(C)$. The ratio $\sigma(C)/s$ can be regarded as a “compression coefficient” for the given s -term sequence. The mathematical expectation of this ratio

$$\mu_s = \frac{\sum_C p(C)\sigma(C)}{s}$$

(where the summation is over all sequences of length s) is the “average compression” for sequences of length s ; finally, the quantity

$$\mu = \overline{\lim_{s \rightarrow \infty}} \mu_s$$

which we shall call the *compression coefficient* (of a given text for a given code) can clearly serve to measure in a natural way the compression of the given uncoded text by the given means of coding. In addition, we note that in all cases actually encountered the quantity μ_s approaches a definite limit as $s \rightarrow \infty$.

We are interested in the smallest value of the compression coefficient that can be achieved by coding when the text has a given statistical structure. Naturally, at the same time we are interested in how to construct the corresponding “optimum” code. A complete answer to these questions is contained in the following remarkably simple theorem:

Theorem 4.

If the entropy of the given text is H , then the greatest lower bound of the compression coefficient μ for all possible codes is $H/\lg m$, where m is the number of different symbols of the text.

Thus, to find the greatest lower bound of the possible shortening of a text by coding, there is no need to know in detail the statistical structure of the text; it is sufficient to consider only its entropy and the number of symbols it uses. Since $\lg m$ is the maximum value of H for the given number of symbols, the quantity $H/\lg m$ is sometimes called the “relative entropy” of the given text.

Proof.

We must show that 1) for any code $\mu \geq \frac{H}{\lg m}$ and 2) for arbitrarily small $\eta > 0$, there exists a code for which $\mu < \frac{H+\eta}{\lg m}$.

1) Choose an arbitrary code and let $H' = H - 2\eta$, where $\eta > 0$ is arbitrarily small. We agree to call an s -term sequence of the given text a *special* sequence, if $\sigma(C) < H's/\lg m$. Since the number of different k -term sequences of the coded text is m^k , the number of all special s -term sequences of the text is no greater than

$$m + m^2 + \dots + m^{\left[\frac{H's}{\lg m}\right]} \leq m^{\frac{H's}{\lg m}} \left\{ 1 + \frac{1}{m} + \frac{1}{m^2} + \dots \right\} = \frac{m}{m-1} a^{sH'},$$

where a is the base of the system of logarithms used. We shall denote the sum of the probabilities of all the special sequences by λ_s and show that $\lambda_s \rightarrow 0$ as $s \rightarrow \infty$.

In fact, let $\varepsilon > 0$ be an arbitrarily small constant; then, by Theorem 3, to obtain a sum of probabilities equal to ε , we must take

$$N_s(\varepsilon) > a^{s(H-\eta)} = a^{s(H'+\eta)}$$

of the most probable sequences. But, as we have seen, for sufficiently large s , the number of all special sequences does not exceed

$$\frac{m}{m-1} a^{sH'} < a^{s(H'+\eta)}.$$

Therefore, the sum λ_s of these probabilities is less than ε if s is sufficiently large, i.e., $\lambda_s \rightarrow 0$ as $s \rightarrow \infty$, as was to be shown.

Since for every non-special sequence (C)

$$\sigma(C) \geq \frac{H's}{\lg m},$$

the mathematical expectation of $\sigma(C)$ exceeds $(1-\lambda_s) \frac{H's}{\lg m}$,

and therefore

$$\mu_s \geq (1-\lambda_s) \frac{H'}{\lg m},$$

but since $\lambda_s \rightarrow 0$ as $s \rightarrow \infty$

$$\mu = \overline{\lim}_{s \rightarrow \infty} \mu_s \geq \frac{H'}{\lg m} = \frac{H-2\eta}{\lg m}.$$

Finally, since η is arbitrarily small, we have

$$\mu \geq \frac{H}{\lg m},$$

which proves the first assertion of Theorem 4.

2) Let $\eta > 0$ be arbitrarily small; to prove the second assertion of Theorem 4, we determine directly a code for which $\mu < \frac{H+\eta}{\lg m}$. Let $\delta > 0$ be arbitrarily small. The number of different sequences of length $\frac{s(H+\delta)}{\lg m}$ equals

$$m^{\frac{s(H+\delta)}{\lg m}} = a^{s(H+\delta)}.$$

In addition, let $\varepsilon > 0$ be arbitrarily small: by Theorem 3, the number $N_s(1-\varepsilon)$ of most probable s -term sequences the sum of which has probability greater than or equal to $1-\varepsilon$ is less than $a^{s(H+\delta)}$, if s is sufficiently large. Therefore, all these “high-probability” s -term sequences can be coded using sequences of length $\frac{s(H+\delta)}{\lg m}$, since, as we see, there are enough of the latter to do so. As regards the remaining “low-probability” s -term sequences (with total probability less than ε), we simply code each of them into itself. In order to guarantee uniqueness of decoding, it is sufficient (for example) to put one of the previously unused sequences of length $\frac{s(H+\delta)}{\lg m}$ (but always the same one!) in front of each such s -term sequence of the coded text. For a code chosen in this way, the length $\sigma(C)$ of a coded s -term sequence will be either $\frac{s(H+\delta)}{\lg m}$ or $s + \frac{s(H+\delta)}{\lg m}$, where the first eventuality occurs with a probability less than or equal to unity and the second with a probability less than or equal to ε . Therefore, the mathematical expectation of $\sigma(C)$ does not exceed

$$\frac{s(H+\delta)}{\lg m} + \varepsilon \left[s + \frac{s(H+\delta)}{\lg m} \right] = s \left[\frac{H+\delta}{\lg m} (1+\varepsilon) + \varepsilon \right] < s \frac{H+\eta}{\lg m},$$

if ε and δ are chosen sufficiently small.

Suppose now we have a large text (C) of length $S=ks$, where k is very large. This text can be broken up into k sequences C_1, C_2, \dots, C_k of length s ; correspondingly, the coded text of length $\sigma(C)$ falls into k sequences with corresponding lengths $\sigma(C_1), \sigma(C_2), \dots, \sigma(C_k)$, so that

$$\sigma(C) = \sigma(C_1) + \sigma(C_2) + \cdots + \sigma(C_k).$$

Therefore, the mathematical expectation of $\sigma(C)$ is k times the mathematical expectation of the quantity $\sigma(C_1)$, and so by the foregoing, does not exceed

$$ks\left(\frac{H+\eta}{\lg m}\right) = S \frac{H+\eta}{\lg m}.$$

It follows that

$$\mu_s \leq \frac{H+\eta}{\lg m},$$

and consequently that

$$\mu \leq \frac{H+\eta}{\lg m},$$

which proves the second assertion of Theorem 4.

(Translated by R. A. Silverman)

*On the Fundamental Theorems of
Information Theory*

On the Fundamental Theorems of Information Theory

(*Uspekhi Matematicheskikh Nauk*, vol. XI, no. 1, 1956, pp. 17-75)

INTRODUCTION

Information theory is one of the youngest branches of applied probability theory; it is not yet ten years old. The date of its birth can, with certainty, be considered to be the appearance in 1947-1948 of the by now classical work of Claude Shannon [1]. Rarely does it happen in mathematics that a new discipline achieves the character of a mature and developed scientific theory in the first investigation devoted to it. Such in its time was the case with the theory of integral equations, after the fundamental work of Fredholm; so it was with information theory after the work of Shannon.

From the very beginning, information theory presents mathematics with a whole new set of problems, including some very difficult ones. It is quite natural that Shannon and his first disciples, whose basic goal was to obtain practical results, were not able to pay enough attention to these mathematical difficulties at the beginning. Consequently, at many points of their investigations they were compelled either to be satisfied with reasoning of an inconclusive nature or to limit artificially the set of objects studied (sources, channels, codes, etc.) in order to simplify the proofs. Thus, the whole mass of literature of the first years of information theory, of necessity, bears the imprint of mathematical incompleteness which, in particular, makes it extremely difficult for mathematicians to become acquainted with this new subject. The recently published general textbook on information theory by S. Goldman [2] can serve as a typical example of the style prevalent in this literature.

Investigations, with the aim of setting information theory on a solid mathematical basis have begun to appear only in recent years and, at the present time, are few in number. First of all, we must mention the work of McMillan [3] in which the fundamental concepts of the theory of discrete sources (source, channel, code, etc.) were first given precise mathematical definitions. The most important result of this work must be considered to be the proof of the remarkable theorem that any discrete ergodic source has the property which Shannon attributed to sources of Markov type and which underlies almost all the asymptotic calculations of information theory.* This circumstance permits the whole theory of discrete information to be constructed without being limited, as was Shannon, to Markov type sources. In the rest of his paper McMillan tries to put Shannon's fundamental theorem on channels with noise on a rigorous basis. In doing so, it becomes apparent that the sketchy proof given by Shannon contains gaps which remain even in the case of Markov sources. The elimination of these gaps is begun in McMillan's paper, but is not completed.

Next, it is necessary to mention the work of Feinstein [4]. Like McMillan, Feinstein considers the Shannon theorem on channels with noise to be the pinnacle of the general theory of discrete information and he undertakes to give a mathematically rigorous proof of this theorem. Accepting completely McMillan's mathematical apparatus, he avoids following Shannon's original path and constructs a proof, using the completely new and apparently very fruitful idea of a "distinguishable set of sequences", the principal features of which will be explained below. However, Feinstein carries out the proof in

* Sections 5-8 of this paper are devoted to this theorem.

all details only for the simplest and least practical case, where the successive signals of the source are mutually independent and the channel memory is zero. In the more general case, he indicates only sketchily how the reader is to carry out the necessary reasoning independently. Unfortunately, there remains a whole series of significant difficulties.

As is well known, Shannon formulated his theorem on channels with noise in two different ways. One was in terms of a quantity called *equivocation*, and the other was in terms of the *probability of error*. McMillan's analysis leads to the conclusion that these two formulations are not equivalent, and that the second gives a more exact result than the first. Feinstein's more detailed investigation showed that although the first formulation is implied by the second, a rigorous derivation of this implication is not only non-trivial but fraught with considerable additional difficulties. Since both formulations are equally important in actual content, it is preferable to speak about two Shannon theorems rather than combine them under the same heading.

In this paper I attempt to give a complete, detailed proof of both of these Shannon theorems, assuming any ergodic source and any stationary channel with a finite memory. At the present time, apparently, these are the broadest hypotheses under which the Shannon theorems can be regarded as valid. On the whole, I follow the path indicated in the works of McMillan and Feinstein, deviating from them only in the comparatively few cases when I see a gap in their explanation, or when another explanation seems to me more complete and convincing (and sometimes, more simple).

The first chapter of the paper, which is of purely auxiliary

character, requires special explanation. It is devoted to the derivation of a whole set of unrelated inequalities, each of which is a theorem of elementary probability theory (i.e., pertains only to finite spaces). The reader acquainted with my paper [5] will be able to begin this paper with the second chapter, returning to the first chapter only when references to its results appear in the text. All the following chapters are constructed according to a specific plan, and can not be skipped or read in different order.

The reader will see that the path to the Shannon theorems is long and thorny, but apparently science, at this time, knows no shorter path if we do not want artificial restrictions on the material studied and if we are to avoid making statements which we can not prove.

CHAPTER I.

Elementary Inequalities

#1. Two generalizations of Shannon's inequality

Let A be a finite probability space composed of elementary events A_i with probabilities

$$p(A_i) \quad (1 \leq i \leq n, p(A_i) > 0; \sum_{i=1}^n p(A_i) = 1).$$

The quantity*

$$H(A) = - \sum_{i=1}^n p(A_i) \lg p(A_i)$$

is called the *entropy* of the space A . The significance of this quantity as a measure of uncertainty or as the amount of information contained in the space A was illustrated by us in detail in [5]. Important properties of the entropy are enumerated there.

Let us now consider, along with A , another finite space B , with the elementary events B_k and the distribution $p(B_k)$ ($1 \leq k \leq m, p(B_k) > 0; \sum_{k=1}^m p(B_k) = 1$). The events A_i and B_k of the spaces A and B can be dependent. The events $A_i B_k$ with the probabilities $p(A_i B_k)$ can be regarded as elementary events of a new finite space which we shall designate by AB (or BA), and which we shall call the *product* of the spaces A and B . The entropy of this space is

$$H(AB) = - \sum_{i=1}^n \sum_{k=1}^m p(A_i B_k) \lg p(A_i B_k).$$

If it is known that the event A_i occurred, then the events B_k

* In this paper all logarithms are to the base 2.

of the space B have the new probabilities

$$p_{A_i}(B_k) = \frac{p(A_i B_k)}{p(A_i)} \quad (k=1, 2, \dots, m)$$

instead of the previous $p(B_k)$. Correspondingly, the previous entropy of the space B

$$H(B) = - \sum_{k=1}^m p(B_k) \lg p(B_k)$$

is replaced by the new quantity

$$H_{A_i}(B) = - \sum_{k=1}^m p_{A_i}(B_k) \lg p_{A_i}(B_k),$$

which, naturally, we shall regard as the conditional entropy of the space B under the assumption that the event A_i occurred in the space A . A specific value of $H_{A_i}(B)$ corresponds to each of the events A_i of the space A , so that $H_{A_i}(B)$ can be regarded as a random variable defined on the space A . The mathematical expectation of this random variable

$$H_A(B) = \sum_{i=1}^m p(A_i) H_{A_i}(B)$$

is the conditional entropy of the space B averaged over the space A . It indicates how much information is contained on the average in the space B , if it is known which of the events of the space A actually occurred. In my paper [5], it is shown that

$$(1.1) \quad H(AB) = H(A) + H_A(B),$$

a relation which is very natural from the standpoint of the real meaning of the quantities in it; in the special case where the spaces A and B are (mutually) independent, we have $H_A(B) = H(B)$, and therefore

$$H(AB) = H(A) + H(B).$$

Shannon's fundamental inequality (also introduced in my paper) is especially important for the purposes of this section. It states that for any finite spaces A and B

$$(1.2) \quad H_A(B) \leq H(B),$$

the real meaning of which is that, on the average, the amount of uncertainty in the space B can either decrease or remain the same, if it is known which event occurred in some other space A . (The uncertainty of a situation can not be increased as a result of obtaining any additional information.) It follows from (1.1) and (1.2) that

$$(1.3) \quad H(AB) \leq H(A) + H(B),$$

an inequality which is easily generalized to the case of the product of any number of spaces; thus

$$H(ABC) \leq H(A) + H(B) + H(C),$$

etc. The inequality (1.2) can be generalized in various directions; we now prove two such generalizations, which will be needed subsequently.

Let A and B be any two finite spaces. We keep all the notation introduced above. In expanded form, (1.2) becomes

$$\sum_{i=1}^n p(A_i) \sum_{k=1}^m p_{A_i}(B_k) \lg p_{A_i}(B_k) \geq \sum_{k=1}^m p(B_k) \lg p(B_k).$$

For later use, it is important to show that this inequality remains valid when we sum both sides not over all, but only over certain values of the subscript k (but, of course, over the same values in both sides of the inequality). In other words, this inequality does not depend on whether or not the events B_k form a "complete system".

Lemma 1.1.

$$(1.4) \quad \sum_{i=1}^n p(A_i) \sum_k^* p_{A_i}(B_k) \lg p_{A_i}(B_k) \geq \sum_k^* p(B_k) \lg p(B_k),$$

where \sum_k^* denotes summation over certain values of the subscript k (not necessarily all, but the same values in both sides of the inequality).

Remark.

In the special case where the summation is carried out over all k ($1 \leq k \leq m$), the left and right sides of (1.4) become $-H_A(B)$ and $-H(B)$, respectively, so that (1.4) agrees with (1.2). Therefore, (1.2) is a special case of Lemma 1.1, and hence is proved when it is.

Proof.

The function $f(x) = x \lg x$ is convex for $x > 0$; consequently, for $x_i > 0$, $\lambda_i \geq 0$, $\sum_{i=1}^n \lambda_i = 1$, the following inequality holds (see [5], [6])

$$(1.5) \quad \sum_{i=1}^n \lambda_i f(x_i) \geq f\left(\sum_{i=1}^n \lambda_i x_i\right).$$

Putting $\lambda_i = p(A_i)$, $x_i = p_{A_i}(B_k)$, we find

$$\begin{aligned} \sum_{i=1}^n p(A_i) \sum_k^* p_{A_i}(B_k) \lg p_{A_i}(B_k) &= \sum_k^* \left\{ \sum_{i=1}^n p(A_i) f[p_{A_i}(B_k)] \right\} \\ &\geq \sum_k^* f \left\{ \sum_{i=1}^n p(A_i) p_{A_i}(B_k) \right\} = \sum_k^* f[p(B_k)] = \sum_k^* p(B_k) \lg p(B_k), \text{ Q.E.D.} \end{aligned}$$

Now we generalize the inequality (1.2) in another direction.

Lemma 1.2.

For any three finite spaces A , B , C

$$H_{AB}(C) \leq H_B(C).$$

Remark.

In the special case where the space B consists of one event with probability 1, the space AB coincides with A , and $H_B(C) = H(C)$. For this special case, the statement of Lemma 1.2 becomes

$$H_A(C) \leq H(C),$$

and is equivalent to the inequality (1.2), which is therefore a special case of Lemma 1.1.

Proof.

Let a given event B_k occur in the space B ; then all the events $A_i C_l$ in the product space AC have the probabilities $p_{B_k}(A_i C_l) = q(A_i C_l)$; in just the same way the space A becomes the space A' with the probabilities $p_{B_k}(A_i) = q(A_i)$ and the space C becomes C' with the probabilities $p_{B_k}(C_l) = q(C_l)$. Hence, according to inequality (1.2)

$$(1.6) \quad H_{A'}(C') \leq H(C').$$

But

$$(1.7) \quad H(C') = - \sum_l q(C_l) \lg q(C_l) = - \sum_l p_{B_k}(C_l) \lg p_{B_k}(C_l) = H_{B_k}(C).$$

On the other hand

$$H_{A'}(C') = - \sum_i q(A_i) \sum_l q_{A_i}(C_l) \lg q_{A_i}(C_l),$$

where

$$q_{A_i}(C_l) = \frac{q(A_i C_l)}{q(A_i)} = \frac{p_{B_k}(A_i C_l)}{p_{B_k}(A_i)} = p_{B_k A_i}(C_l);$$

consequently

$$(1.8) \quad H_{A'}(C') = - \sum_i p_{B_k}(A_i) \sum_l p_{B_k A_i}(C_l) \lg p_{B_k A_i}(C_l)$$

$$= \sum_i p_{B_k}(A_i) H_{B_k A_i}(C).$$

Substituting (1.7) and (1.8) into (1.6), we find

$$\sum_i p_{B_k}(A_i) H_{B_k A_i}(C) \leq H_{B_k}(C).$$

Multiplying both sides of this inequality by $p(B_k)$ and then summing over all k , we obtain

$$\sum_{i, k} p(B_k) p_{B_k}(A_i) H_{B_k A_i}(C) \leq \sum_k p(B_k) H_{B_k}(C),$$

or

$$\sum_{i, k} p(A_i B_k) H_{A_i B_k}(C) \leq H_B(C),$$

or, finally

$$H_{AB}(C) \leq H_B(C), \quad \text{Q. E. D.}$$

#2. Three Inequalities of Feinstein [4]

We again consider two finite spaces A and B and their product AB . Let Z be some set of events $A_i B_k$ of the space AB , and let U_0 be some set of events A_i of the space A ; let $\delta_1 > 0$, $\delta_2 > 0$, and

$$p(Z) > 1 - \delta_1; \quad p(U_0) > 1 - \delta_2.$$

Let us denote by I'_i ($1 \leq i \leq n$) the set of events B_k of the space B for which $A_i B_k$ does not belong to the set Z . Finally, let U_1 be the set of events $A_i \in U_0$ for which $p_{A_i}(I'_i) \leq \alpha$. Then we have

Lemma 2.1.

$$p(U_1) > 1 - \delta_2 - \frac{\delta_1}{\alpha}.$$

Proof.

Let U_2 be the set of events A_i for which $p_{A_i}(\Gamma_i) > \alpha$, so that

$$(2.1) \quad U_1 = U_0 - U_0 U_2.$$

If $A_i \in U_2$, then

$$p(A_i \Gamma_i) = p(A_i) p_{A_i}(\Gamma_i) > \alpha p(A_i),$$

which means that

$$p\left(\sum_{A_i \in U_2} A_i \Gamma_i\right) = \sum_{A_i \in U_2} p(A_i \Gamma_i) > \alpha \sum_{A_i \in U_2} p(A_i) = \alpha p(U_2).$$

On the other hand, since all the events $A_i \Gamma_i$ are incompatible with the event Z , the probability of which exceeds $1 - \delta_1$, then

$$p\left(\sum_{A_i \in U_2} A_i \Gamma_i\right) \leq 1 - p(Z) < \delta_1,$$

and we find that $\alpha p(U_2) < \delta_1$, or $p(U_2) < \delta_1/\alpha$. It follows, *a fortiori*, that

$$p(U_0 U_2) < \frac{\delta_1}{\alpha};$$

hence, by (2.1)

$$p(U_1) = p(U_0) - p(U_0 U_2) > 1 - \delta_2 - \delta_1/\alpha, \quad \text{Q. E. D.}$$

For a given subscript k ($1 \leq k \leq m$), we now designate by i_k the value of the subscript i ($1 \leq i \leq n$), for which the probability $p(A_i B_k)$ assumes its greatest value (in the space AB); if there is more than one such i value, we take any one as i_k . Thus A_{i_k} is the event of the space A which is most probable for a given event B_k of the space B . Obviously, the sum

$$P = \sum_{k=1}^m \sum_{\substack{i=1 \\ i \neq i_k}}^n p(A_i B_k)$$

is the probability (in the space AB) of the appearance of a pair of events $A_i B_k$ such that A_i is not the event of the space A which is most probable for a given event B_k of the space B . Clearly, it is possible to write

$$P = p(i \neq i_k).$$

Lemma 2.2.

If for a given ε ($0 < \varepsilon < 1$), a set \mathcal{A}_i of events B_k can be associated with each A_i ($1 \leq i \leq n$) such that

$$1) \quad p(\mathcal{A}_i \mathcal{A}_j) = 0 \quad (i \neq j)$$

$$2) \quad p_{A_i}(\mathcal{A}_i) > 1 - \varepsilon \quad (1 \leq i \leq n)$$

then $P \leq \varepsilon$.

Proof.

Obviously, we have

$$(2.2) \quad P = 1 - \sum_{k=1}^m p(A_{i_k} B_k); \quad 1 - P = \sum_{k=1}^m p(A_{i_k} B_k).$$

Let us denote by \mathcal{A}_0 the set of events B_k (if such exist) which are not in any of the sets \mathcal{A}_i ($1 \leq i \leq n$); then, clearly, the range of summation in the last sum can be expanded into the parts $\mathcal{A}_0, \mathcal{A}_1, \dots, \mathcal{A}_n$, and therefore

$$1 - P = \sum_{i=1}^n \sum_{B_k \in \mathcal{A}_i} p(A_{i_k} B_k) + \sum_{B_k \in \mathcal{A}_0} p(A_{i_k} B_k) \leq \sum_{i=1}^n \sum_{B_k \in \mathcal{A}_i} p(A_{i_k} B_k).$$

According to the definition of the subscript i_k , the right side of this inequality can only be decreased if we replace the subscript i_k in each term of the sum by any other subscript from 1 to n . Thus, in particular

$$\begin{aligned} 1 - P &\geq \sum_{i=1}^n \sum_{B_k \in \mathcal{A}_i} p(A_i B_k) = \sum_{i=1}^n p(A_i \mathcal{A}_i) = \sum_{i=1}^n p(A_i) p_{A_i}(\mathcal{A}_i) \\ &> (1 - \varepsilon) \sum_{i=1}^n p(A_i) = 1 - \varepsilon, \quad \text{Q. E. D.} \end{aligned}$$

As before, let n denote the number of elementary event A_i of the space A .

Lemma 2.3.

For $n > 1$

$$H_B(A) \leq P \lg (n-1) - P \lg P - (1-P) \lg (1-P).$$

Proof.

As above, for brevity, we write $f(x)$ for $x \lg x$. We have

$$H_B(A) = - \sum_k p(B_k) \sum_i f[p_{B_k}(A_i)] = H_1 + H_2$$

where

$$H_1 = - \sum_k p(B_k) f[p_{B_k}(A_{i_k})],$$

$$H_2 = - \sum_k p(B_k) \sum_{i \neq i_k} f[p_{B_k}(A_i)].$$

Putting $\lambda_k = p(B_k)$, $x_k = p_{B_k}(A_{i_k})$ in the inequality (1.5), we find by (2.2) that

$$(2.3) \quad \begin{aligned} H_1 &= - \sum_k \lambda_k f(x_k) \leq - f\left(\sum_k \lambda_k x_k\right) = - f\left(\sum_k p(B_k) p_{B_k}(A_{i_k})\right) \\ &= - f\left(\sum_k p(B_k A_{i_k})\right) = - f(1-P) = -(1-P) \lg (1-P). \end{aligned}$$

A similar application of the inequality (1.5) yields because of (2.2)

$$(2.4) \quad \begin{aligned} - \sum_k p(B_k) f[1 - p_{B_k}(A_{i_k})] &\leq - f\left(\sum_k p(B_k)[1 - p_{B_k}(A_{i_k})]\right) \\ &= - f\left(1 - \sum_k p(B_k A_{i_k})\right) = - f(P) = - P \lg P. \end{aligned}$$

Furthermore, since

$$1 - p_{B_k}(A_{i_k}) = \sum_{i \neq i_k} p_{B_k}(A_i),$$

then for any fixed k ($1 \leq k \leq m$)

$$\begin{aligned} f[1 - p_{B_k}(A_{i_k})] &= \sum_{i \neq i_k} p_{B_k}(A_i) \lg \sum_{i \neq i_k} p_{B_k}(A_i) \\ &= (n-1)f\left(\frac{\sum_{i \neq i_k} p_{B_k}(A_i)}{n-1}\right) + \lg (n-1) \sum_{i \neq i_k} p_{B_k}(A_i). \end{aligned}$$

We again use (1.5) to estimate the first term on the right side by putting this time $\lambda_i = \frac{1}{n-1}$, $x_i = p_{B_k}(A_i)$ ($1 \leq i \leq n, i \neq i_k$). This gives

$$f\left(\frac{\sum_{i \neq i_k} p_{B_k}(A_i)}{n-1}\right) \leq \sum_{i \neq i_k} \frac{1}{n-1} f[p_{B_k}(A_i)],$$

and, therefore

$$f[1 - p_{B_k}(A_{i_k})] \leq \sum_{i \neq i_k} f[p_{B_k}(A_i)] + \lg (n-1) \sum_{i \neq i_k} p_{B_k}(A_i).$$

Multiplying all the terms of this inequality by $p(B_k)$ and summing over k from i to m , we find by (2.2) that

$$\begin{aligned} (2.5) \quad \sum_k p(B_k) f[1 - p_{B_k}(A_{i_k})] &\leq \sum_k p(B_k) \sum_{i \neq i_k} f[p_{B_k}(A_i)] \\ &\quad + P \lg (n-1) = -H_2 + P \lg (n-1). \end{aligned}$$

Finally, combining (2.3), (2.4), and (2.5) term by term, we find

$$H_1 \leq -(1-P) \lg (1-P) - P \lg P - H_2 + P \lg (n-1),$$

from which

$$H_B(A) = H_1 + H_2 \leq P \lg (n-1) - P \lg P - (1-P) \lg (1-P),$$

Q. E. D.

CHAPTER II.

Ergodic Sources

#3. Concept of source. Stationarity. Entropy

In statistical communication theory, the output of every information source is regarded as a random process. The statistical structure of this process constitutes the mathematical definition of the given source. In this paper, we shall deal exclusively with *discrete* sources (processes with discrete time); the output of such a source is a sequence of random quantities (or events). Consequently, we must understand the definition of a source to be the complete probabilistic characterization of such a sequence.

Underlying the definition of every source is the set A of symbols used by it, which we call its *alphabet*, and which we always assume to be *finite*. The separate symbols of this alphabet, are called its *letters*. Let us consider a sequence of letters, infinite on both sides.

$$(3.1) \quad x = (\dots, x_{-1}, x_0, x_1, x_2, \dots),$$

which represents a possible “life history” of the given source. We shall regard it as an elementary event in a certain (infinite) probability space, a space the specification of which characterizes the sequence (3.1) as a random process. The set of all sequences (3.1) (i. e., the set of all elementary events of the given space) will be denoted by A^I . Any subset of the set A^I represents an event of our space, and conversely. Thus, for example, the event “the source emits the letter α at time t

and the letter β at time u " is the set of all sequences x of the form (3.1) for which

$$x_t = \alpha, \quad x_u = \beta.$$

Generally, if t_1, t_2, \dots, t_n are any integers and $\alpha_1, \alpha_2, \dots, \alpha_n$ are any letters of the alphabet A , then the event "the source emits the letter α_i at time t_i ($1 \leq i \leq n$)" is the set of all x for which

$$x_{t_i} = \alpha_i \quad (1 \leq i \leq n).$$

Subsequently, we shall call such a set of elementary events x a *cylinder set*, or briefly a *cylinder*. As is well known, it is sufficient to know the probability $\mu(Z)$ of all cylinders Z to define the sequence (3.1) as a random process. Let us consider the set of all cylinders of the given alphabet A and its *Borel extension* F_A , i.e., the intersection of all the Borel fields which contain all the cylinders of the alphabet A . Then, giving the probabilities $\mu(Z)$ of all cylinders Z uniquely determines the probability $\mu(S)$ of any set $S \in F_A$ of elementary events x . Thus a complete description of the source as a random process is achieved by specifying 1) an alphabet A , and 2) a probability measure $\mu(S)$ defined for all $S \in F_A$. In particular, we always have $\mu(A^I) = 1$. Since the alphabet A and the probability measure μ completely characterize the statistical nature of the source, we can denote the source by the symbol $[A, \mu]$.

In his basic work [1], Shannon considered only sources with the character of stationary Markov chains; the characterization of such sources is achieved by more elementary means. The general concept of source just given is due to McMillan [3].

Consider any given sequence x of the type (3.1) of letters from the alphabet A , and denote by Tx the sequence

$$Tx = (\dots, x'_{-1}, x'_0, x'_1, x'_2, \dots),$$

where $x'_k = x_{k+1}$ ($-\infty < k < +\infty$) (so that the operator T denotes the "shift" by one time unit). If S is any set of elements x , then the set TS is the set of all Tx for which $x \in S$. (In other words, the relations $x \in S$ and $Tx \in TS$ are equivalent.) It is easy to see that if $S \in F_A$ we have $TS \in F_A$; it is also obvious that the operator T maps the set A' into itself

$$TA' = A'.$$

If

$$\mu(TS) = \mu(S)$$

for any set $S \in F_A$, then the source is called *stationary*. Evidently, by the stationarity of a source is meant the time invariance of the probability regime of its output. All the sources considered below will be assumed to be stationary.

From the information theory viewpoint, the most important characteristic of every source is the rate at which it emits information, i.e., the average amount of information given by one emitted symbol. We now show how to arrive at an exact definition of this quantity. Consider a sequence of n successive symbols emitted by a given source; let this sequence be $x_t, x_{t+1}, \dots, x_{t+n-1}$. If the source alphabet A contains a letters, then the number of such different n -term sequences is obviously a^n . Every such sequence C is a cylinder in the space A' (i.e., the set of all $x \in A'$ for which the x_t, \dots, x_{t+n-1} assume the fixed values characterizing the given sequence), and therefore has a definite probability $\mu(C)$. Thus, the set of all possible n -term sequences of the type described represents a finite probability space consisting of a^n elementary events C with probabilities

$\mu(C)$. In chapter I we agreed to measure the amount of information contained in such a space by its entropy

$$H_n = - \sum_C \mu(C) \lg \mu(C).$$

If the given source is stationary (as we shall always assume), then the probabilities $\mu(C)$, and therefore the entropy H_n of an n -term sequence do not depend on the "initial moment" t , and are uniquely determined by the nature of the source and by the number n . Thus we can say that the sequence of n symbols emitted by the source gives a well defined amount of information H_n , which depends only on n and on the nature of the source, so that, on the average, the amount of information per symbol emitted by the source is H_n/n . Consequently, it is natural to agree to call the quantity

$$H = \lim_{n \rightarrow \infty} \frac{H_n}{n}$$

the *source entropy*, i. e., the average amount of information conveyed by one emitted symbol (if, of course, the limit exists). Clearly, the entropy H as thus defined, depends only on the nature of the source (i. e., on the alphabet A and on the probability distribution μ). The fact that the entropy really exists for every stationary source is the first fundamental theorem of the general theory of discrete sources. We turn to its proof.*

The space A_{n+m} of sequences of length $n+m$ (where n and m are positive integers) can be regarded as the product of the space A_n of sequences of length n and the space A_m of sequences of length m (see #1). Hence, according to #1

* As far as I know, the first proof was given by McMillan [3]. However, McMillan had in mind a broader aim, so that his proof is considerably more formidable than the one given below.

$$H(A_{n+m}) = H(A_n) + H_{A_n}(A_m); \quad H_{A_n}(A_m) \leq H(A_m),$$

from which

$$H(A_n) \leq H(A_{n+m}) \leq H(A_n) + H(A_m).$$

In our new notation, this can be written as

$$H_n \leq H_{n+m} \leq H_n + H_m.$$

In particular, the first of these inequalities yields (for $m=1$)

$$(3.2) \quad H_n \leq H_{n+1},$$

and the second is easily extended to any number of terms and, in particular, gives for any integral k

$$(3.3) \quad H_{kn} \leq kH_n.$$

Setting $n=1$ in (3.3), we find that for any $k \geq 1$.

$$H_k \leq kH_1,$$

which shows that

$$a = \liminf_{n \rightarrow \infty} \frac{H_n}{n} < +\infty.$$

Now let $\varepsilon > 0$ be given arbitrarily, and let the subscript q be chosen such that

$$\frac{H_q}{q} < a + \varepsilon.$$

For any $n > q$, we determine an integer $k > 1$ such that

$$(k-1)q < n \leq kq.$$

Then, because of (3.2)

$$H_n \leq H_{kq}.$$

Consequently, by (3.3)

$$\frac{H_n}{n} \leq \frac{H_{kq}}{(k-1)q} \leq \frac{k}{k-1} \frac{H_q}{q} < \frac{k}{k-1} (a + \varepsilon),$$

and therefore, for sufficiently large n (and hence for appropriately large k)

$$a - \varepsilon < \frac{H_n}{n} < \frac{k}{k-1} (a + \varepsilon) < a + 2\varepsilon.$$

But since ε is arbitrarily small

$$\lim_{n \rightarrow \infty} \frac{H_n}{n} = a, \quad \text{Q. E. D.}$$

#4. Ergodic Sources

The set S of elements $x \in A^I$ is called *invariant* if $TS = S$, i.e., if the “shift operator” T carries the set into itself. The set A^I is always invariant. For any $x \in A^I$, the set of elements $\dots, T^{-1}x, x, Tx, T^2x, \dots$ is always an invariant set. The source $[A, \mu]$ is called *ergodic* if the probability $\mu(S)$ of every invariant set $S \in F_A$ is either 0 or 1. The ergodic property is very important in the study of the statistical structure of the source. This results from the following considerations. Each (numerically valued) function $f(x)$ of the elementary event $x \in A^I$ can be regarded as a random variable defined on the space A^I , and conversely. If the abstract Lebesgue integral

$$(4.1) \quad \int_{A^I} |f(x)| d\mu(x) < +\infty,$$

then this random variable has the mathematical expectation

$$Mf(x) = \int_{A^I} f(x) d\mu(x).$$

The well-known “ergodic theorem” of Birkhoff states that

for any stationary source $[A, \mu]$ and for every summable function $f(x)$ (i.e., satisfying the requirement (4.1)) the limit

$$\lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=0}^{n-1} f(T^k x) = h(x)$$

exists almost everywhere (i.e., with probability 1) where the function $h(x)$ is invariant, i.e., $h(Tx) = h(x)$ for all x for which $h(x)$ exists. If the source $[A, \mu]$ is ergodic, then almost everywhere (with probability 1)

$$h(x) = Mf(x);$$

thus, in the case of an ergodic source, the ratio

$$\frac{\sum_{k=0}^{n-1} f(T^k x)}{n}$$

approaches the mathematical expectation of $f(x)$ as $n \rightarrow \infty$, for almost all x .

Now let $g_S(x)$ be the characteristic function of the set $S \in F_A$ (i.e., $g_S(x) = 1$ ($x \in S$) and $g_S(x) = 0$ ($x \notin S$)). Obviously, $g_S(x)$ is summable and $Mg_S(x) = \mu(S)$. The sum $\sum_{k=0}^{n-1} g_S(T^k x)$ is the number of terms of the series $x, Tx, \dots, T^{n-1}x$ which belong to the set S . Let us denote this number by $\varphi_{n, S}(x)$. Then the Birkhoff theorem states that in the case of an ergodic source, we have for almost all x

$$\lim_{n \rightarrow \infty} \frac{\varphi_{n, S}(x)}{n} = \mu(S).$$

Thus, in this case, the proportion of the terms in the sequence

$$x, Tx, T^2x, \dots$$

which are elements of the set is just the probability of the set S , for almost all such sequences.

Let us agree to say that the source $[A, \mu]$ reflects the set $S \in F_A$, if almost everywhere (with probability 1)

$$\lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=0}^{n-1} g_S(T^k x) = \mu(S).$$

We have just seen that an ergodic source reflects any set $S \in F_A$. It is easy to convince oneself that the converse holds: If the stationary source $[A, \mu]$ reflects any set $S \in F_A$, then the source is ergodic. Indeed, if the $[A, \mu]$ source were not ergodic, then there would exist an invariant set $S \in F_A$ for which $0 < \mu(S) < 1$. Because of the invariance of S , we would have $Tx \in S$, $T^2x \in S$, ... for any $x \in S$ and, therefore $g_S(T^k x) = 1$ ($k = 0, 1, 2, \dots$). Consequently

$$\lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=0}^{n-1} g_S(T^k x) = 1 \neq \mu(S)$$

for any $x \in S$. Since $\mu(S) > 0$, then the set S is not reflected by the source $[A, \mu]$, Q. E. D.

A somewhat stronger theorem will be needed below: *In order for a given stationary source to be ergodic, it is sufficient for it to reflect all cylinders of the space A^I .* To see this, it is enough to show that the sets $S \in F_A$, which are reflected by a given source, generate a Borel field, for, if this field contains all cylinders of the space A^I , it contains the whole field F_A , by the very definition of the latter. But this means that the given source reflects any set $S \in F_A$ and, as we have just seen, this is sufficient for the source to be ergodic.

Let us denote by G the class of all sets $S \in F_A$ which are reflected by the given source; in our special case, G contains all cylinders of the space A^I . If the sets S_1, S_2, \dots, S_k are non-overlapping, then obviously

$$g_{S_1 + \dots + S_k}(x) = g_{S_1}(x) + \dots + g_{S_k}(x),$$

whence it follows at once that $S_1 \in G, \dots, S_k \in G$ implies $S_1 + \dots + S_k \in G$. Furthermore, if $S_2 \subset S_1$, then

$$g_{S_1 - S_2}(x) = g_{S_1}(x) - g_{S_2}(x),$$

whence it is easy to see that $S_1 - S_2 \in G$ follows from $S_1 \in G$ and $S_2 \in G$. In order to convince ourselves that G is a Borel field, it remains only to show that the rule which we established above for a finite sum of sets $S_1 + \dots + S_k$ remains valid for an infinite sum $\sum_{k=1}^{\infty} S_k$ of non-overlapping sets. For this purpose we need the following auxiliary proposition.

Lemma 4.1.

For any $\varepsilon > 0$, no matter how small, there exists a $\delta > 0$ such that if the set $U \in F_A$ has probability less than δ , then $\mu(S_\varepsilon) < \varepsilon$, where S_ε is the set of values of x for which $h_U(x) > \varepsilon$.

Proof.

For $x \in S_\varepsilon$

$$\lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=0}^{n-1} g_U(T^k x) = h_U(x) > \varepsilon.$$

Hence, if n is sufficiently large, we have

$$(4.2) \quad \sum_{k=0}^{n-1} g_U(T^k x) > \frac{\varepsilon}{2} n$$

for all points x of a set S'_ε with probability

$$\mu(S'_\varepsilon) \geq \frac{1}{2} \mu(S_\varepsilon).$$

From (4.2) we find by integrating over x

$$(4.3) \quad \sum_{k=0}^{n-1} \int_{S'_\varepsilon} g_U(T^k x) d\mu(x) \geq \frac{\varepsilon n}{2} \mu(S'_\varepsilon) \geq \frac{\varepsilon n}{4} \mu(S_\varepsilon).$$

But, on the other hand, because of the stationarity of the given source

$$\int_{S'_\varepsilon} g_U(T^k x) d\mu(x) = \int_{S'_\varepsilon} g_U(T^k x) d\mu(T^k x) \leq \mu(U) \quad (0 \leq k < n),$$

and (4.3) gives

$$\frac{\varepsilon n}{4} \mu(S_\varepsilon) \leq n\mu(U).$$

If we put $\delta = \frac{\varepsilon^2}{4}$, then the inequality

$$\mu(S_\varepsilon) < \varepsilon$$

follows from $\mu(U) < \delta$, which proves the lemma.

Now, let $S_i \in G$ ($i = 1, 2, \dots$), $\sum_{i=1}^{\infty} S_i = S$ and let the sets S_i be non-overlapping so that the series

$$\sum_{i=1}^{\infty} \mu(S_i) = \mu(S)$$

converges. We put $\sum_{i>N} S_i = U_N$, so that

$$\mu(U_N) \rightarrow 0 \quad (N \rightarrow \infty).$$

Since $S = S_1 + \dots + S_N + U_N$, then

$$g_S(x) = g_{S_1}(x) + \dots + g_{S_N}(x) + g_{U_N}(x).$$

and therefore

$$h_S(x) = h_{S_1}(x) + \dots + h_{S_N}(x) + h_{U_N}(x).$$

But we have $h_{S_i}(x) = \mu(S_i)$ ($1 \leq i \leq N$) with probability 1. Therefore, with probability 1, we have

$$h_S(x) = \mu\left(\sum_{i=1}^N S_i\right) + h_{U_N}(x).$$

But by lemma 4.1, for sufficiently large N and with probability arbitrarily close to unity, we have

$$h_{U_N}(x) < \varepsilon,$$

where $\varepsilon > 0$ is arbitrarily small. Since $h_s(x)$ is independent of N and ε , then with probability 1

$$h_s(x) = \mu(S),$$

and therefore $S \in G$. Thus, G is a Borel field, a fact which, as we have seen, completes the proof that every source which reflects all the cylinders of the space A' is ergodic.

#5. The E property. McMillan's theorem.

If the alphabet A of a given source contains a letters, then the number of different " n -term sequences"

$$x_t, x_{t+1}, \dots, x_{t+n-1}$$

which can be emitted by the source is a^n . As already stated in #3, these sequences C can be regarded as the elementary events of a finite probability space. The probabilities $\mu(C)$ of these elementary events are determined by specifying the source itself, since every sequence C is a cylinder of the space A' and has the definite probability $\mu(C)$ (which in the case of a stationary source is independent of t). Thus, the different sequences C can also be regarded as compound events (cylinders) of the infinite space A' . As long as n remains constant, the first point of view is preferable, since it is considerably simpler than the second. But if n varies in our considerations, then the first point of view becomes inconvenient, since it forces us to consider a separate space for every value of n ; in such cases it is usually more advantageous to use the second point of view, since it

allows us to consider all the events as occurring in one and the same space A' .

As we have already noted in #3, every (numerical valued) function of the letters $x_t, x_{t+1}, \dots, x_{t+n-1}$ can be regarded as a random variable on the space A' of the given source, since the letters $x_t, x_{t+1}, \dots, x_{t+n-1}$ as well as the value of the given function are uniquely determined by specifying the elementary event of the space A' ($x = \dots, x_{-1}, x_0, x_1, \dots$). In particular, $-(1/n) \lg \mu(C)$ is such a random variable, where C is a random sequence $x_t, x_{t+1}, \dots, x_{t+n-1}$ with given n and t . (Of course, $\lg \mu(C)$ is independent of t if the given source is stationary.) Clearly, this random variable assumes the same value for all x having the identical sequence $C = x_t, x_{t+1}, \dots, x_{t+n-1}$ (i. e., belonging to the cylinder corresponding to this sequence). Consequently, the mathematical expectation of this quantity can easily be found by elementary means, i. e., by multiplying its value on each cylinder C by the probability $\mu(C)$ of the cylinder and adding all the products. This gives

$$M\left(-\frac{1}{n} \lg \mu(C)\right) = -\frac{1}{n} \sum_C \mu(C) \lg \mu(C).$$

Here, we recognize the sum $-\sum_C \mu(C) \lg \mu(C)$ (see #3) as the entropy of n -term sequences from the given source, which we denoted by H_n . Thus we find

$$M\left[-\frac{1}{n} \lg \mu(C)\right] = \frac{H_n}{n}.$$

Assuming the given source to be stationary, we set $t=0$, so that hereafter C denotes the sequence x_0, x_1, \dots, x_{n-1} . The random variable $-(1/n) \lg \mu(C)$ is then a function of x and n , which we denote by $f_n(x)$; thus

$$Mf_n(x) = \frac{H_n}{n}.$$

In #3 we showed that for any stationary source the ratio H_n/n approached a definite limit as $n \rightarrow \infty$, which we agreed to call the entropy (per letter) of the given source. Thus we find for any stationary source

$$Mf_n(x) \rightarrow H \quad (n \rightarrow \infty),$$

i.e., the mathematical expectation of the random variable $f_n(x) = -(1/n) \lg \mu(C)$ approaches the entropy of the given source as $n \rightarrow \infty$.

The fact of primary significance for all of information theory is that a considerably stronger statement holds under certain simplifying assumptions about the nature of the source: Not only does the mathematical expectation of $f_n(x)$ approach H as a limit as $n \rightarrow \infty$, but $f_n(x)$ itself converges in probability to H as $n \rightarrow \infty$. This means that, for arbitrarily small $\epsilon > 0$ and $\delta > 0$, the probability of the inequality $|f_n(x) - H| > \epsilon$ is less than δ for sufficiently large n . This can be said still more descriptively as follows, by recalling the definition of $f_n(x)$. For arbitrarily small $\epsilon > 0$ and $\delta > 0$, and for sufficiently large n , all the n -term sequences C in the output of the given source can be separated into two groups, such that

- 1) For every sequence C of the first group

$$\left| \frac{\lg \mu(C)}{n} + H \right| < \epsilon.$$

- 2) The sum of the probabilities of all the sequences of the second group is less than δ .

Let us agree to call the first group the “high probability group”; and the second group the “low probability group”.

The high probability group is characterized by the fact that $(1/n) \lg \mu(C)$ is close to $-H$ for all its sequences C , so that the probability $\mu(C)$ of such a sequence is approximately 2^{-nH} . Hence, all the sequences of the high probability group have approximately the same probability 2^{-nH} , which means that the number of sequences in this group is approximately 2^{nH} . If we recall that the number of all n -term sequences is $a^n = 2^{n \lg a}$ and that H is always $\leq \lg a$, then we see that, generally speaking (more exactly, with the exception of the case $H = \lg a$), for large n the high probability group contains only a negligibly small share of all the n -term sequences from the source. On the contrary, the overwhelming majority of such sequences fall into the low probability group.

We shall call the source property just described the E property. As already mentioned, it is of fundamental significance in information theory. Therefore it is important to find the broadest possible class of sources possessing this property. For sources such that each letter is statistically independent of all the preceding letters, the E property is an almost immediate consequence of the law of large numbers, and always holds; however, in practice, we seldom encounter such sources. As Shannon showed (see [1], also [5], Theorem 3), all sources of the simple ergodic Markov chain type also possess the E property, and this proof is easily carried over to the case of compound Markov chains of any order. Finally, in 1953, McMillan succeeded in proving that any ergodic source possesses the E property. This important theorem permitted for the first time the construction of a mathematical basis for the general theory of information with sufficiently broad assumptions on the statistical nature of the transmitted information. We shall give

a detailed proof of McMillan's theorem in subsequent sections. In doing so, we shall refer to certain "ergodic theorems" of a general character, which, like Birkhoff's theorem (already cited more than once), we must assume to be known to the reader.

#6. The martingale concept. Doob's theorem.

In recent years, the concept of *martingale*, introduced by Doob, has been useful in various problems of probability theory. Here, we must become acquainted with this concept in a rather limited form corresponding to our needs.

Let

$$(6.1) \quad \xi_1, \xi_2, \dots, \xi_m, \dots$$

be a sequence of random variables defined on the space of elementary events $x \in A^I$. (The sequence of functions $f_n(x)$ which we considered in the previous section is such a sequence.) In general, the ξ_m are mutually dependent, and, for $m > 1$, we can speak of the conditional mathematical expectation of ξ_m for given values of ξ_1, \dots, ξ_{m-1} . Let us agree to denote the conditional mathematical expectation of ξ_m for $\xi_1 = a_1, \xi_2 = a_2, \dots, \xi_{m-1} = a_{m-1}$ by $M_{a_1 a_2 \dots a_{m-1}}(\xi_m)$.

The sequence (6.1) is called a *martingale* if for any $m > 1$ and any a_i

$$(6.2) \quad M_{a_1 \dots a_{m-1}}(\xi_m) = a_{m-1}.$$

We call the martingale (6.1) bounded if the random variables ξ_m are all bounded ($|\xi_m| < C$ for any x and m). In what follows, we deal only with bounded martingales, so that the question of the existence of the mathematical expectations which we need causes no difficulty.

By the “sequence a_m, \dots, a_n ”, where $m \leq n$, we mean the set (cylinder) of those $x \in A$ for which $\xi_i = a_i$ ($m \leq i \leq n$). Let A_{m-1} be the sequence a_1, \dots, a_{m-1} ; then

$$M_{a_1 \dots a_{m-1}}(\xi_m) = \frac{1}{\mu(A_{m-1})} \int_{A_{m-1}} \xi_m(x) d\mu(x),$$

and the characteristic requirement (6.2), defining the martingale, can be written as

$$\int_{A_{m-1}} \xi_m(x) d\mu(x) = a_{m-1} \mu(A_{m-1}).$$

The fundamental theorem of Doob [7], which we need, is that *every bounded martingale converges with probability 1* (almost everywhere in A'). We need two auxiliary propositions in order to prove it.

Lemma 6.1.

Let $\xi_1, \xi_2, \dots, \xi_m, \dots$ be a bounded martingale, and let A_j be one of the sequences a_1, a_2, \dots, a_j , where $m > j$. Then

$$(6.3) \quad \int_{A_j} \xi_m d\mu(x) = \int_{A_j} \xi_j d\mu(x) = a_j \mu(A_j).$$

(In other words, the conditional mathematical expectation of ξ_m is a_j , for given a_1, \dots, a_j).

Proof.

For $m = j+1$, (6.3) follows from the definition of a martingale. Consequently, let $m > j+1$. Then

$$\int_{A_j} \xi_m d\mu(x) = \sum_s \int_{A_j C_s} \xi_m d\mu(x)$$

where the summation is over all sequences C_s of the form a_{j+1}, \dots, a_{m-1} . For any sequence C_s , the intersection $A_j C_s$ is some

sequence a_1, \dots, a_{m-1} , so that, by the martingale definition

$$\int_{A_j C_s} \xi_m d\mu(x) = a_{m-1} \mu(A_j C_s) = \int_{A_j C_s} \xi_{m-1} d\mu(x).$$

(since $\xi_{m-1} = a_{m-1}$ for any $x \in A_j C_s$). Summing this equality over all the sequences C_s , we find (for $m > j+1$)

$$\int_{A_j} \xi_m d\mu(x) = \int_{A_j} \xi_{m-1} d\mu(x).$$

Clearly, the successive application of this recurrence relation leads to (6.3) and Lemma 6.1 is proved.

Lemma 6.2.

Let $n > m > 0$, let N be a set of sequences a_1, \dots, a_m , and let Λ be the set of all sequences a_m, \dots, a_n for which

$$\min_{m \leq j \leq n} a_j \leq k,$$

where k is any given real number. Then, for $r > n$

$$\int_{\Lambda N} \xi_r(x) d\mu(x) \leq k \mu(\Lambda N).$$

Proof.

We denote by Λ_j ($m \leq j \leq n$) the set of all sequences a_m, \dots, a_n for which

$$a_m > k, \dots, a_{j-1} > k, a_j \leq k.$$

Clearly, the sets Λ_j ($m \leq j \leq n$) are non-overlapping and have the set Λ as their union. It is also clear that each Λ_j is a certain set of sequences a_1, \dots, a_j (i. e., whether $x \in \Lambda_j$ or not is completely determined by the values of the quantities $\xi_1(x) = a_1, \dots, \xi_j(x) = a_j$). We have

$$(6.4) \quad \int_{\Lambda N} \xi_r d\mu(x) = \sum_{j=m}^n \int_{\Lambda_j N} \xi_r d\mu(x).$$

But each $\Lambda_j N$ is, evidently, a certain set of sequences $C_s^{(j)}$ of the type a_1, \dots, a_j , $\Lambda_j N = \sum_s C_s^{(j)}$, so that by Lemma 6.1 for $r > n$

$$\int_{\Lambda_j N} \xi_r d\mu(x) = \sum_s \int_{C_s^{(j)}} \xi_r d\mu(x) = \sum_s \int_{C_s^{(j)}} \xi_j d\mu(x) = \int_{\Lambda_j N} \xi_j d\mu(x).$$

Since $\xi_j = a_j \leq k$ on the set Λ_j , then

$$\int_{\Lambda_j N} \xi_r d\mu(x) \leq k \mu(\Lambda_j N) \quad (m \leq j \leq n),$$

and, therefore, by (6.4)

$$\int_{\Lambda N} \xi_r d\mu(x) \leq k \mu(\Lambda N), \quad \text{Q. E. D.}$$

Remark.

It is clear that

$$\int_{\Lambda N} \xi_r d\mu(x) \geq k \mu(\Lambda N)$$

can be proved in a completely analogous way if Λ is the set of all sequences a_m, \dots, a_n for which

$$\max_{m \leq j \leq n} a_j \geq k.$$

Doob's Theorem.

Every bounded martingale (6.1) converges with probability 1 (i.e., almost everywhere in A').

Proof.

Suppose the theorem is incorrect. Then, real numbers k_1 and k_2 can be found such that

$$\liminf_{n \rightarrow \infty} \xi_n < k_1 < k_2 < \limsup_{n \rightarrow \infty} \xi_n$$

on a set Λ of positive probability. We put $\mu(\Lambda) = \eta > 0$. Clearly, we can select a positive integer n_1 so large that

$$\max_{1 \leq j \leq n_1} \xi_j \geq k_2$$

on a set Λ_1 for which

$$\mu(\Lambda \Lambda_1) > \eta \left(1 - \frac{1}{3}\right).$$

Furthermore, if n_2 is large enough, we have

$$\min_{n_1 < j \leq n_2} \xi_j \leq k_1$$

on a set Λ_2 for which

$$\mu(\Lambda \Lambda_2) > \eta \left(1 - \frac{1}{3^2}\right);$$

again, if n_3 is large enough

$$\max_{n_2 < j \leq n_3} \xi_j \geq k_2$$

on a set Λ_3 for which

$$\mu(\Lambda \Lambda_3) > \eta \left(1 - \frac{1}{3^3}\right),$$

etc. We continue this alternating process indefinitely. We put

$$\Lambda_1 \Lambda_2 \cdots \Lambda_k = M_k \quad (k=1,2,\dots).$$

For any $k \geq 1$, we have

$$\begin{aligned} \mu(M_k) &\geq \mu(\Lambda M_k) = \mu(\Lambda) - \mu(\Lambda \bar{M}_k) = \eta - \mu \left\{ \sum_{i=1}^k (\Lambda - \Lambda \Lambda_i) \right\} \\ &\geq \eta - \sum_{i=1}^k \mu(\Lambda - \Lambda \Lambda_i) = \eta - \sum_{i=1}^k [\mu(\Lambda) - \mu(\Lambda \Lambda_i)] \\ &\geq \eta - \sum_{i=1}^k \left[\eta - \eta \left(1 - \frac{1}{3^i}\right) \right] = \eta - \eta \sum_{i=1}^k \frac{1}{3^i} > \eta - \frac{\eta}{2} = \frac{\eta}{2}. \end{aligned}$$

Furthermore, we have for any $r \geq 1$

$$M_{2r} = \Lambda_{2r} M_{2r-1}.$$

It is clear that in this expression M_{2r-1} is some set of sequences $a_1, a_2, \dots, a_{n_{2r-1}}$ and Λ_{2r} is the set of sequences $a_{n_{2r-1}}, \dots, a_{n_{2r}}$ for which

$$\min_{n_{2r-1} \leq j \leq n_{2r}} a_j \leq k_1.$$

Thus, we see that if we put $m = n_{2r-1}$, $n = n_{2r}$, $N = M_{2r-1}$, $\Lambda = \Lambda_{2r}$, $k = k_1$ in Lemma 6.2, then all the premises of this lemma are fulfilled. Therefore, if $m > n_{2r}$, Lemma 6.2 gives

$$\int_{M_{2r}} \xi_m d\mu(x) \leq k_1 \mu(M_{2r}).$$

In a completely analogous way (see the remark after the proof of Lemma 6.2), we have

$$\int_{M_{2r-1}} \xi_m d\mu(x) \geq k_2 \mu(M_{2r-1}).$$

Therefore, putting $M_{2r-1} - M_{2r} = Q_r$, we find

$$\begin{aligned} \int_{Q_r} \xi_m d\mu(x) &= \int_{M_{2r-1}} \xi_m d\mu(x) - \int_{M_{2r}} \xi_m d\mu(x) \geq k_2 \mu(M_{2r-1}) - k_1 \mu(M_{2r}) \\ &= (k_2 - k_1) \mu(M_{2r-1}) + k_1 [\mu(M_{2r-1}) - \mu(M_{2r})] > \\ &> (k_2 - k_1) \frac{\eta}{2} + k_1 \mu(Q_r). \end{aligned}$$

Since the sets Q_r are non-overlapping, the series $\sum_{n=1}^{\infty} \mu(Q_r)$ converges, and $\mu(Q_r) \rightarrow 0$ as $r \rightarrow \infty$. But since the left side of the last inequality does not exceed $C\mu(Q_r)$ because $|\xi_m(x)| < C$, then both the left side and the second term of the right side of this inequality are infinitesimally small as $r \rightarrow \infty$, whereas the first term of the right side is a positive constant. Thus we arrive at a contradiction which proves Doob's theorem.

#7. Auxiliary propositions

We have already remarked repeatedly (##3,5) that every quantity which can be uniquely determined by the sequence x_i, \dots, x_{i+n-1} of letters from the alphabet A can be regarded as a function of the elementary event

$$x = \dots, x_{-1}, x_0, x_1, \dots$$

of the space A^I or, equivalently, as a random variable on this space. In particular, in #5 we defined such a function, viz. $f_n(x) = -(1/n) \lg \mu(C)$, where C is the sequence x_0, x_1, \dots, x_{n-1} . Now we must become acquainted with some other functions of the same kind. Let us agree to denote the random sequence x_{-n}, \dots, x_{-1} by C_n , and the sequence $x_{-n}, \dots, x_{-1}, x_0$ with one extra letter by $C_n + x_0$. Each of these sequences is an event (cylinder) of the space A^I , and every quantity which is uniquely determined by the letters $x_{-n}, \dots, x_{-1}, x_0$ is a random variable on this space or, equivalently, a function of the elementary event x . In particular, the ratio

$$p_n(x) = \frac{\mu(C_n + x_0)}{\mu(C_n)}$$

is such a quantity, and can obviously be regarded as the conditional probability of the appearance of the letter x_0 after the appearance of the sequence $C_n = x_{-n}, \dots, x_{-1}$. In addition, let us put $p_0(x) = \mu(x_0)$.

Let α be any fixed letter of the alphabet A , and put

$$p_n(x, \alpha) = \frac{\mu(C_n + \alpha)}{\mu(C_n)}$$

This quantity (the probability of the appearance of the letter α after the sequence C_n) differs from $p_n(x)$ in that in $p_n(x)$

the letter x_0 depends (just as do the letters x_{-n}, \dots, x_{-1}) on x , whereas the letter α which replaces it in $P_n(x, \alpha)$ is fixed and independent of x .

Lemma 7.1.

The sequence $p_n(x, \alpha)$ ($n=0, 1, \dots$) is a martingale.

Proof.

For brevity, let us put $p_n(x, \alpha) = \xi_n$. Consider any fixed sequence of $n-1$ letters $a_{-1}, \dots, a_{-(n-1)}$ and denote by B_{n-1} the cylinder $x_{-1} = a_{-1}, \dots, x_{-(n-1)} = a_{-(n-1)}$ of the space A^I . Furthermore, let Γ_β denote the cylinder $x_{-n} = \beta$, where β is any letter of the alphabet A . Since $\sum_\beta \Gamma_\beta = A^I$, then

$$\int_{B_{n-1}} \xi_n d\mu(x) = \sum_\beta \int_{B_{n-1} \Gamma_\beta} \xi_n d\mu(x),$$

but, for $x \in B_{n-1} \Gamma_\beta$, we have

$$\xi_n = \frac{\mu(B_{n-1} \Gamma_\beta + \alpha)}{\mu(B_{n-1} \Gamma_\beta)}.$$

Therefore

$$\int_{B_{n-1}} \xi_n d\mu(x) = \sum_\beta \mu(B_{n-1} \Gamma_\beta + \alpha) = \mu(B_{n-1} + \alpha).$$

But according to the definition of the random variable

$$\xi_{n-1} = \frac{\mu(C_{n-1} + \alpha)}{\mu(C_{n-1})},$$

its value $[\xi_{n-1}]_{B_{n-1}}$ at $C_{n-1} = B_{n-1}$ (i.e., when the random cylinder C_{n-1} coincides with the cylinder B_{n-1} which we chose) equals:

$$[\xi_{n-1}]_{B_{n-1}} = \frac{\mu(B_{n-1} + \alpha)}{\mu(B_{n-1})}.$$

Therefore

$$(7.1) \quad \int_{B_{n-1}} \xi_n d\mu(x) = [\xi_{n-1}]_{B_{n-1}} \mu(B_{n-1}).$$

Now, let K_{n-1} be the set of all x for which ξ_1, \dots, ξ_{n-1} take any given system of values $\xi_i = \pi_i$ ($1 \leq i \leq n-1$). Since the numbers π_i ($1 \leq i \leq n-1$) are uniquely determined by the selection of the cylinder B_{n-1} , the set K_{n-1} is the union of several cylinders B_{n-1} , where, of course, $[\xi_i]_{B_{n-1}} = [\xi_i]_{K_{n-1}} = \pi_i$ ($1 \leq i \leq n-1$) for all the cylinders B_{n-1} in K_{n-1} . Thus it follows from (7.1) that

$$\int_{K_{n-1}} \xi_n d\mu(x) = \sum_{B_{n-1} \subset K_{n-1}} \int_{B_{n-1}} \xi_n d\mu(x) = \sum_{B_{n-1} \subset K_{n-1}} \pi_{n-1} \mu(B_{n-1}) = \pi_{n-1} \mu(K_{n-1}).$$

But this means that the sequence $\xi_n = p_n(x, \alpha)$ ($n \geq 0$) is a martingale, which proves Lemma 7.1.

Lemma 7.2.

The sequence $p_n(x)$ ($n=0, 1, \dots$) converges almost everywhere.

Proof.

It is clear that for every fixed $x \in A'$, the quantity $p_n(x)$ coincides with one of the $p_n(x, \alpha)$, where α runs through all the letters of the alphabet A , and where α is the same for all n (but different for the different x). Consequently, for any n, m and for any $x \in A'$.

$$|p_n(x) - p_m(x)| \leq \sum_{\alpha} |p_n(x, \alpha) - p_m(x, \alpha)|.$$

But the martingale $p_n(x, \alpha)$, being obviously bounded, converges almost everywhere, by Doob's theorem (#6), no matter what the letter α is, and, therefore, the right side of the last inequality approaches zero when n and m increase without limit; clearly this proves Lemma 7.2.

The definition which we gave of the function $p_n(x)$ assumes that $\mu(C_n) > 0$; the x for which $\mu(C_n) = 0$ are obviously a set of probability 0, and we exclude them from consideration from

the outset by keeping only the x for which $\mu(C_n) > 0$ for any $n \geq 0$. Thus, the sequence $p_n(x)$ is defined almost everywhere in A^I . But then the sequence of functions

$$g_n(x) = -\lg p_n(x) \quad (n=0,1,\dots)$$

will be defined almost everywhere, where the value $g_n(x) = +\infty$ [for $p_n(x) = 0$] is not excluded for these functions.

Lemma 7.3.

Let $E_{n,k}$ ($n \geq 0, k \geq 0$) denote the set of all x for which

$$k \leq g_n(x) < k+1;$$

then

$$(7.2) \quad \int_{E_{n,k}} g_n(x) d\mu(x) \leq a(k+1)2^{-k},$$

where a is the number of letters in the alphabet A .

Proof.

We define the cylinders B_n as in the proof of Lemma 7.1, and let Z_α denote the cylinder $x_0 = \alpha$, where α is any letter of the alphabet A . For $x \in B_n Z_\alpha$, we have:

$$g_n(x) = -\lg \frac{\mu(B_n + \alpha)}{\mu(B_n)} = -\lg \frac{\mu(B_n Z_\alpha)}{\mu(B_n)},$$

so that the value of $g_n(x)$ is uniquely determined by assigning the cylinder B_n and the letter α . It is clear that

$$B_n E_{n,k} = \sum_{\alpha}^* B_n Z_\alpha,$$

where the \sum_{α}^* denotes summation over those letters α for which $k \leq g_n(x) < k+1$ for $x \in B_n Z_\alpha$. Therefore

$$(7.3) \quad \int_{B_n E_{n,k}} g_n(x) d\mu(x) = \sum_{\alpha}^* \int_{B_n Z_\alpha} g_n(x) d\mu(x).$$

But in any of the integrals of the right side we have

$$k+1 > g_n(x) = -\lg \frac{\mu(B_n Z_a)}{\mu(B_n)} \geq k,$$

so that

$$\mu(B_n Z_a) \leq 2^{-k} \mu(B_n),$$

and (7.3) gives

$$\int_{B_n E_{n,k}} g_n(x) d\mu(x) \leq \sum_{\sigma}^* (k+1) \mu(B_n Z_a) \leq a(k+1) 2^{-k} \mu(B_n).$$

Finally, summation over all possible cylinders B_n gives

$$\int_{E_{n,k}} g_n(x) d\mu(x) \leq a(k+1) 2^{-k}$$

and Lemma 7.3 is proved. As we shall see immediately, the fact that this estimate is uniform with respect to n is important. In particular, the next proposition (uniform summability of the functions $g_n(x)$) is an immediate consequence of Lemma 7.3.

Lemma 7.4.

Given $L > 0$, let $A_{n,L}$ be the set of $x \in A^I$ for which $g_n(x) > L$. Then, given any $\varepsilon > 0$, an $L_0 = L_0(\varepsilon)$ can be found such that

$$\int_{A_{n,L}} g_n(x) d\mu(x) < \varepsilon,$$

for $L \geq L_0$, and for any $n \geq 1$.

Another important consequence of Lemma 7.3 is that the absolute continuity of integrals of the functions $g_n(x)$ is also uniform with respect to n . Thus we have

Lemma 7.5.

Given any $\varepsilon > 0$, a $\delta > 0$ can be found such that if $E \in F_A$ and $\mu(E) < \delta$, then

$$\int_E g_n(x) d\mu(x) < \varepsilon, \quad (n=1,2,\dots).$$

Proof.

By Lemma 7.4, an $L=L(\varepsilon)$ can be found such that

$$\int_{A_{n,L}} g_n(x) d\mu(x) < \varepsilon, \quad (n=1,2,\dots).$$

Put $\delta = \frac{\varepsilon}{L}$ and let $\mu(E) < \delta$. Then

$$\begin{aligned} \int_E g_n(x) d\mu(x) &= \int_{E \setminus A_{n,L}} g_n(x) d\mu(x) + \int_{\overline{E \setminus A_{n,L}}} g_n(x) d\mu(x) \\ &\leq \int_{A_{n,L}} g_n(x) d\mu(x) + L\mu(E) < 2\varepsilon, \end{aligned}$$

and Lemma 7.5 is proved. A consequence is that almost everywhere on A' we have $g_n(x) < +\infty$ for any $n \geq 1$.

Now let us put

$$\lim_{n \rightarrow \infty} g_n(x) = g(x).$$

This limit exists almost everywhere, because of Lemma 7.2, if we permit the value $+\infty$ for $g(x)$.

Lemma 7.6.

The function $g(x)$ is summable over A' . (In particular, this means that it can take the value $+\infty$ only on a set of probability 0.)

Proof.

For any $L > 0$ and any positive function $f(x)$, we put

$$f^L(x) = \min [L, f(x)].$$

Then obviously it follows from $g_n(x) \rightarrow g(x)$ that $g_n^L(x) \rightarrow g^L(x)$

$(n \rightarrow \infty)$. But since the functions $g_n^L(x)$ ($n = 1, 2, \dots$) are uniformly bounded, we find, using the known properties of the Lebesgue integral and Lemma 7.3, that

$$\begin{aligned} \int_{A^I} g^L(x) d\mu(x) &= \lim_{n \rightarrow \infty} \int_{A^I} g_n^L(x) d\mu \leq \limsup_{n \rightarrow \infty} \int_{A^I} g_n(x) d\mu(x) \\ &= \limsup_{n \rightarrow \infty} \sum_{k=0}^{\infty} \int_{E_{n,k}} g_n(x) d\mu(x) < a \sum_{k=0}^{\infty} (k+1) 2^{-k}. \end{aligned}$$

Since the right side is independent of L , Lemma 7.6 is proved.

Lemma 7.7.

$$\int_{A^I} |g_n(x) - g(x)| d\mu(x) \rightarrow 0 \quad (n \rightarrow \infty).$$

Proof.

Let $\varepsilon > 0$ be arbitrary. Let us denote by E_n the set of $x \in A^I$ for which $|g_n(x) - g(x)| > \varepsilon$. Then

$$\begin{aligned} \int_{A^I} |g_n(x) - g(x)| d\mu(x) &= \int_{E_n} |g_n(x) - g(x)| d\mu(x) + \int_{\overline{E_n}} |g_n(x) - g(x)| d\mu(x) \\ &\leq \int_{E_n} g_n(x) d\mu(x) + \int_{\overline{E_n}} g(x) d\mu(x) + \varepsilon. \end{aligned}$$

Since $g_n(x) \rightarrow g(x)$ almost everywhere, $\mu(E_n) \rightarrow 0$ ($n \rightarrow \infty$); consequently, the first integral on the right side is infinitesimally small as $n \rightarrow \infty$ by Lemma 7.5, and the same is true of the second integral by Lemma 7.6. This proves Lemma 7.7.

#8. Proof of McMillan's theorem

McMillan's theorem states (see #5) that as $n \rightarrow \infty$ the function $f_n(x)$ converges in probability to the number H , the entropy of the given source. Here

$$f_n(x) = -\frac{1}{n} \lg \mu(C),$$

where C is the random sequence (cylinder) x_0, x_1, \dots, x_{n-1} (which is, of course, uniquely determined by giving $x = \dots, x_{-1}, x_0, x_1, \dots$). In order to use the results of #7, we must first of all relate the functions $f_n(x)$ to the functions $g_n(x)$, which we studied in #7. This relation is established by the following proposition.

Lemma 8.1.

For all $x \in A^I$ and $n \geq 1$

$$f_n(x) = \frac{1}{n} \sum_{k=0}^{n-1} g_k(T^k x).$$

Proof.

Since we shall now deal with sequences x_r, \dots, x_{r+s} for various values of r and s , we must introduce a more extensive notation. We denote the probability of such a sequence by $\mu[x_r, \dots, x_{r+s}]$ so that, for example

$$f_n(x) = -\frac{1}{n} \lg \mu[x_0, \dots, x_{n-1}].$$

Similarly, the function $p_n(x)$, which we introduced in #7, can be written as

$$p_n(x) = \frac{\mu[x_{-n}, \dots, x_0]}{\mu[x_{-n}, \dots, x_{-1}]}.$$

From this it is clear that for $k \geq 0$

$$p_n(T^k x) = \frac{\mu[x_{k-n}, \dots, x_k]}{\mu[x_{k-n}, \dots, x_{k-1}]},$$

and, in particular

$$p_k(T^k x) = \frac{\mu[x_0, \dots, x_k]}{\mu[x_0, \dots, x_{k-1}]}.$$

This equality holds for $k \geq 1$. Since $p_0(x) = \mu[x_0]$ by definition, then

$$p_0(T^0x) = p_0(x) = \mu[x_0],$$

so that

$$\prod_{k=0}^{n-1} p_k(T^k x) = \mu[x_0] \cdot \frac{\mu[x_0, x_1]}{\mu[x_0]} \cdots \frac{\mu[x_0, \dots, x_{n-1}]}{\mu[x_0, \dots, x_{n-2}]} = \mu[x_0, \dots, x_{n-1}].$$

Taking the logarithm of this equality yields

$$\sum_{k=0}^{n-1} g_k(T^k x) = -\log \mu[x_0, \dots, x_{n-1}] = n f_n(x),$$

and Lemma 8.1 is proved.

McMillan's Theorem.

For any stationary source the sequence $f_n(x)$ converges in the L^1 -mean (and therefore also in probability) to some invariant function $h(x)$. In the case of an ergodic source, $h(x)$ coincides almost everywhere in A^I with the entropy H of the source.

Remark.

$f_n(x)$ is said to approach $h(x)$ in the L^1 -mean if

$$\int_{A^I} |f_n(x) - h(x)| d\mu(x) \rightarrow 0 \quad (n \rightarrow \infty);$$

it is well-known that this implies that $f_n(x)$ converges to $h(x)$ in probability. As already noted, the function $h(x)$ is called invariant if $h(Tx) = h(x)$ ($x \in A^I$).

Proof.

One of the familiar forms of the ergodic theorem ([8], p. 31, Satz 10; [9], equation (2.42)) states that for any function $g(x)$ which is summable over A^I , the quantity

$$\frac{1}{n} \sum_{k=0}^{n-1} g(T^k x)$$

approaches an invariant function $h(x)$ in the L^1 -mean as $n \rightarrow \infty$. In particular, this holds for our function $g(x)$ because of Lemma 7.6. But by Lemma 8.1 we have

$$\begin{aligned} \int_{A^I} |f_n(x) - h(x)| d\mu(x) &= \int_{A^I} \left| \frac{1}{n} \sum_{k=0}^{n-1} g_k(T^k x) - h(x) \right| d\mu(x) \\ &\leq \int_{A^I} \left| \frac{1}{n} \sum_{k=0}^{n-1} [g_k(T^k x) - g(T^k x)] \right| d\mu(x) + \int_{A^I} \left| \frac{1}{n} \sum_{k=0}^{n-1} g(T^k x) - h(x) \right| d\mu(x) \\ &\leq \frac{1}{n} \sum_{k=0}^{n-1} \int_{A^I} |g_k(x) - g(x)| d\mu(x) + \int_{A^I} \left| \frac{1}{n} \sum_{k=0}^{n-1} g(T^k x) - h(x) \right| d\mu(x), \end{aligned}$$

where in transforming the first term we use the stationarity of the source. ($d\mu(x)$ is replaced by $d\mu(T^k x)$, and then $T^k x$ is introduced as a new variable of integration.)

Consider each of the two terms on the right side separately. In the first term, the summands with increasing index approach zero by Lemma 7.7. Therefore, their average value, i.e., the whole first term on the right side, approaches zero as $n \rightarrow \infty$. The same is true for the second term by the very definition of the function $h(x)$. Thus we find

$$\int_{A^I} |f_n(x) - h(x)| d\mu(x) \rightarrow 0 \quad (n \rightarrow \infty),$$

which proves the first part of McMillan's theorem.

In the case of an ergodic source, the ergodic theorem cited above states that the function $h(x)$ is almost everywhere a constant h . Thus, to prove the second part of McMillan's theorem, we must show that $h=H$. But from

$$\int_{A^I} |f_n(x) - h| d\mu(x) \rightarrow 0 \quad (n \rightarrow \infty),$$

it follows that

$$\lim_{n \rightarrow \infty} \int_{A^I} f_n(x) d\mu(x) = \int_{A^I} h d\mu(x) = h.$$

The integral on the left side is the mathematical expectation of $f_n(x)$ which, as we saw in #5, approaches H as $n \rightarrow \infty$. This means that $h = H$, which completes the proof of McMillan's theorem.

In particular, therefore, we have proved that every ergodic source has the E property: *For arbitrarily small $\varepsilon > 0$ and $\delta > 0$, and sufficiently large n , all the a^n n -term sequences of the source output are divided into two groups, a high probability group, such that $|(1/n) \lg \mu(C) + H| < \varepsilon$ for each of its sequences, and a low probability group, such that the sum of the probabilities of its sequences is less than δ .*

CHAPTER III.

Channels and the sources driving them

#9. Concept of channel. Noise. Stationarity. Anticipation and memory

From a physical point of view, by a source is meant an apparatus which emits signals. The medium over which the signal is *transmitted*, is called a channel. The concept of channel, as well as the concept of source, plays a fundamental role in information theory. Just as in Chapter II we started the theory of sources by giving their precise mathematical characterization, we must now do the same thing for channels. We saw that the elements which can be used to characterize a source mathematically are its alphabet A , the probability space A^I with elementary events x , and the probability measure $\mu(S)$ defined on the sets $S \in F_A$. What mathematical elements can be used to characterize a given channel?

First of all, we must have a list of the signals which the channel in question can transmit. We shall always assume that this list is finite, and we shall call it the *input alphabet* (or the *alphabet at the input*) of the channel. The signals in this alphabet are called its *letters*. In general, the signals that emerge from a channel have an entirely different nature from the transmitted signals. Therefore, in addition to the input alphabet, we must know the *output alphabet* (or *alphabet at the output*) of the given channel, i.e., the list of signals (*letters*) which the channel can emit. We shall also always assume that the output alphabet is finite.

If every transmitted signal a (a letter of the input alphabet A) gives at the output a unique letter $b=b(a)$ of the output alphabet B , the channel is called a *noiseless channel*. In general, the presence of interference (noise) causes *different* letters $b \in B$ to be obtained at the output in different cases when the *same* letter $a \in A$ is transmitted. Since it is natural to regard the interference (noise) as a random phenomenon, and it governs which letter $b \in B$ appears at the channel output when a given signal $a \in A$ is transmitted, then we can speak of the *probability* of obtaining the letter $b \in B$ at the channel output, given that the letter $a \in A$ was transmitted. In many cases, this probability depends not only on a but also on the sequence of signals transmitted earlier, and this dependence must often be taken into account. Therefore, at the outset it is expedient to give the most general form to the phenomenon described. As in Chapter II, we shall consider the set A' of all sequences

$$x = \dots, x_{-1}, x_0, x_1, \dots \quad (x_i \in A; -\infty < i < +\infty).$$

To every such sequence of transmitted signals corresponds a sequence

$$y = \dots, y_{-1}, y_0, y_1, \dots \quad (y_i \in B; -\infty < i < +\infty)$$

of received (output) signals. The probability that y_n will be a given letter $b \in B$ must be regarded, in the general case, as depending on the whole set of signals x_i sent, i.e., as a function of $x \in A'$. We can denote this probability by $\nu_x(y_n=b)$. We can also say that $\nu_x(y_n=b)$ is the conditional probability (for a given sequence $x \in A'$ of transmitted signals) that the sequence $y \in B'$ of received signals will belong to the cylinder $y_n=b$. It is clear that in order to characterize the channel completely we must know such conditional probabilities not only for the

simplest cylinders of the type $y_n = b$, but also for sets $S \subset B'$ of more complex structure. It is natural to require that such probabilities $\nu_x(S)$ be assigned for all cylinders $S \subset B'$; in this way, $\nu_x(S)$ will be automatically defined for all sets S in the Borel field F_B generated by the set of these cylinders.

Thus, we shall consider a channel to be specified if we know the following three elements: 1) the input alphabet A , 2) the output alphabet B , 3) the probability $\nu_x(S)$ that the y received when a given x is transmitted belongs to the set $S \in F_B$. (This probability must be given for any $x \in A'$ and any $S \in F_B$.) We shall denote the channel specified by these elements by $[A, \nu_x, B]$. Evidently, $\nu_x(S)$ must be understood to be a one-parameter (with parameter x) family of probability measures defined on the space of elementary events B' . We shall call the channel $[A, \nu_x, B]$ *stationary* if, for all $x \in A'$ and $S \in F_B$

$$\nu_{Tx}(TS) = \nu_x(S),$$

where T is the same shift operator with which we were concerned in the preceding chapter.

In the majority of applications, the probability $\nu_x(y_n = b)$ that y_n coincides with a given letter $b \in B$ does not depend on all the letters $x = \dots, x_{-1}, x_0, x_1, \dots$ of the transmitted message, but only on those with indices rather close to n . First of all, we shall always assume that the distribution of y_n is independent of the transmitted signals that are transmitted after x_n , i.e., is independent of x_k for $k > n$; this means that $\nu_x(y_n = b)$ has the same value for all transmitted messages x for which the signals \dots, x_{n-1}, x_n are identical. In this case, we speak of a *channel without anticipation*. As regards the signals x_{n-1}, x_{n-2}, \dots preceding x_n , in the majority of cases only a limited number

of them (e.g., $x_{n-1}, x_{n-2}, \dots, x_{n-m}$) can influence the distribution of y_n . This means that the probability $\nu_x(y_n=b)$ is the same for all x with identical $x_{n-m}, \dots, x_{n-1}, x_n$. In this case, we speak of a *channel with a finite memory*. The smallest number m for which the above holds, is called the *memory* of the channel; in particular, the distribution of y_n for a channel without memory ($m=0$) depends only on x_n .

#10. Connection of the channel to the source

Suppose we have a channel $[A, \nu_x, B]$ and a source $[A, \mu]$ for which the alphabet A coincides with the input alphabet of the channel. Then, clearly, the channel $[A, \nu_x, B]$ can be used directly to transmit the output of the source $[A, \mu]$. Each symbol x_n emitted by the source is one of the letters of the alphabet A , and therefore can act as an input to the channel $[A, \nu_x, B]$. Then at the channel output we obtain a letter y_n of the alphabet B . When the letters from some message

$$x = \dots, x_{-1}, x_0, x_1, \dots$$

from the given source are fed into the channel one at a time, we obtain at the output the corresponding sequence

$$y = \dots, y_{-1}, y_0, y_1, \dots$$

of letters from the alphabet B . In the general case, the distribution law of every letter y_n of this sequence depends on the entire sequence x , and is determined by the probability measure ν_x . For a channel without anticipation, this distribution law depends only on the symbols \dots, x_{n-1}, x_n ; if, in addition, the channel has finite memory m , then the distribution of y_n depends only on x_{n-m}, \dots, x_n .

This whole situation is described by saying that the source $[A, \mu]$ feeds (drives) the channel $[A, \nu_x, B]$. From the probabilistic point of view, such a connection of the channel and the source is a phenomenon in which chance intervenes in two ways: 1) the choice of the message $x \in A^I$ is random (this randomness is governed by the distribution μ , and 2) for a given $x \in A^I$, the $y \in B^I$ which is received at the channel output is random, because of the presence of noise. (This randomness is governed by the distribution ν_x .)

Consider now the probability space in which the elementary events are all possible pairs (x, y) , $x \in A^I$, $y \in B^I$. Let C be the set of all pairs (a, b) , where $a \in A$, $b \in B$. We can regard C as a new alphabet; then, it is natural to denote by C^I the set of pairs (x, y) of which we just spoke. Thus the specification of $(x, y) \in C^I$ is equivalent to the specification of $x \in A^I$, $y \in B^I$.

We must now introduce probabilities into the space C^I . Let $M \subset A^I$, $N \subset B^I$, i.e., let M be some set of elements x and N some set of elements y ; indeed, let $M \in F_A$ (so that $\mu(M)$ has a definite value) and let $N \in F_B$. Then it is clear that the direct product $S = M \times N$ of the sets M and N is a set of pairs (x, y) , so that $S \subset C^I$. The probability $\omega(S)$ of this set of the space C^I should naturally be understood to be the probability of the joint event $x \in M$, $y \in N$. But the distribution in the space of elementary events $x \in A^I$ is determined by the μ law, and, for a given x , the distribution in the space of elementary events $y \in B^I$ is determined by the ν_x law. Therefore

$$(10.1) \quad \omega(S) = \omega(M \times N) = \int_M \nu_x(N) d\mu(x).$$

In particular, every cylinder $Z \subset C^I$ is obviously the direct product of a cylinder $Z_1 \subset A^I$ and a cylinder $Z_2 \subset B^I$. Therefore, Eq.

(10.1) tells us the probability of any such cylinder Z . It follows by the general property of Borel sets that the probability $\omega(S)$ is uniquely defined for any set $S \in F_c$.

Thus we see that connecting the channel $[A, \nu_x, B]$ to the source $[A, \mu]$ driving it uniquely determines a new source $[C, \omega]$. The alphabet of this source is the direct product $A \times B$; the set C' of elementary events (x, y) is the direct product $A' \times B'$, and the probability measure $\omega(S)$ is given by (10.1) in the way described above. In what follows, we shall call this new source $[C, \omega]$ the *compound* source constructed by connecting the channel $[A, \nu_x, B]$ to the source $[A, \mu]$. Since we shall deal only with stationary sources and channels, the following almost obvious theorem will be of importance to us.

Theorem.

If the source $[A, \mu]$ and the channel $[A, \nu_x, B]$ are stationary, then the source $[C, \omega]$ is also stationary.

Proof.

Suppose $S \subset C'$ and $S = X \times Y$, $X \in F_A$, $Y \in F_B$. Then, first of all, it is obvious that $TS = TX \times TY$. Therefore, Eq. (10.1) gives

$$\omega(TS) = \int_{x \in TX} \nu_x(TY) d\mu(x),$$

or, after substituting a new variable of integration

$$\omega(TS) = \int_{Tz \in TX} \nu_{Tz}(TY) d\mu(Tz).$$

Since the source $[A, \mu]$ is stationary, $d\mu(Tz) = d\mu(z)$, and since the channel $[A, \nu_x, B]$ is stationary, $\nu_{Tz}(TY) = \nu_z(Y)$; therefore

$$\omega(TS) = \int_{z \in X} \nu_z(Y) d\mu(z) = \omega(S).$$

We have established this equality for all $S=X\times Y$, in particular, for all cylinders $S\subset C'$; therefore it is also valid for all $S\in F_c$. This proves the theorem.

Let us put $M=A'$ in Eq. (10.1), while leaving $N\in F_B$ arbitrary. The quantity $\omega(M\times N)$ is then the probability of the joint event $x\in A'$, $y\in N$, and since the first of these two events is sure, $\omega(M\times N)$ is simply the probability $\eta(N)$ of obtaining a sequence y belonging to the set $N\in F_B$ at the channel output. Thus we see that the distribution $\eta(N)$ plays the same role for the space B' as $\mu(M)$ does for the space A' . For $M=A'$, equation (10.1) becomes

$$(10.2) \quad \eta(N) = \omega(A'\times N) = \int_{A'} \nu_x(N) d\mu(x).$$

Thus we can speak of the source $[B, \eta]$ at the channel output. This source, with the sequence

$$y=\dots, y_{-1}, y_0, y_1, \dots$$

of letters from the alphabet B as its output, is uniquely determined (by using (10.2)) by the data of our problem, i.e., by the source A and the channel $[A, \nu_x, B]$. (Of course, the source $[C, \omega]$ considered above is uniquely determined by the same data.) If the source $[A, \mu]$ and the channel $[A, \nu_x, B]$ are stationary, then, as we proved, the source $[C, \omega]$ is also stationary; but then

$$\eta(TN) = \omega(A'\times TN) = \omega(TA'\times TN) = \omega(A'\times N) = \eta(N),$$

and therefore the source $[B, \eta]$ is also stationary.

We proved in #3 that every stationary source has a definite entropy. Therefore, if the source $[A, \mu]$ and the channel $[A, \nu_x, B]$ are stationary, each of the three sources $[A, \mu]$, $[B, \eta]$

and $[C, \omega]$ has a definite entropy. Let us agree to denote these three entropies by $H(X)$, $H(Y)$, and $H(X, Y)$, respectively. We defined the entropy $H(X)$ as follows. Let $H_n(X)$ be the entropy of the finite space the elementary events of which are all the n -term sequences x_0, \dots, x_{n-1} (cylinders) emitted by the source $[A, \mu]$, with corresponding probabilities determined by the distribution μ ; then

$$H(X) = \lim_{n \rightarrow \infty} \frac{1}{n} H_n(X).$$

In complete analogy

$$H(Y) = \lim_{n \rightarrow \infty} \frac{1}{n} H_n(Y),$$

$$H(X, Y) = \lim_{n \rightarrow \infty} \frac{1}{n} H_n(X, Y),$$

where correspondingly $H_n(Y)$ and $H_n(X, Y)$ denote the entropies of the finite spaces of sequences y_0, \dots, y_{n-1} from the source $[B, \eta]$ and $x_0, y_0, \dots, x_{n-1}, y_{n-1}$ from the source $[C, \omega]$.

But giving the “pair sequence” $x_0, y_0, \dots, x_{n-1}, y_{n-1}$ is obviously equivalent to giving the pair of sequences x_0, \dots, x_{n-1} and y_0, \dots, y_{n-1} , so that the space of sequences $x_0, y_0, \dots, x_{n-1}, y_{n-1}$ is the product of the spaces of sequences x_0, \dots, x_{n-1} and y_0, \dots, y_{n-1} . Therefore, as we saw in §1

$$H_n(X, Y) = H_n(X) + H_{nX}(Y) = H_n(Y) + H_{nY}(X),$$

where $H_{nX}(Y)$ denotes the average conditional entropy of the space of sequences y_0, \dots, y_{n-1} for a given sequence x_0, \dots, x_{n-1} , and $H_{nY}(X)$ has an analogous meaning. It follows that

$$H_{nX}(Y) = H_n(X, Y) - H_n(X),$$

$$H_{nY}(X) = H_n(X, Y) - H_n(Y).$$

Dividing all the terms of these equalities by n , and passing to the limit as $n \rightarrow \infty$, we find

$$(10.3) \quad \left\{ \begin{array}{l} \lim_{n \rightarrow \infty} \frac{1}{n} H_{nX}(Y) = H(X, Y) - H(X), \\ \lim_{n \rightarrow \infty} \frac{1}{n} H_{nY}(X) = H(X, Y) - H(Y). \end{array} \right.$$

This means, first of all, that the limits on the left always exist. Let us denote them by $H_X(Y)$ and $H_Y(X)$, respectively. Consider the conditional entropy per symbol of the output of the source $[B, \eta]$ for a given output x of the source $[A, \mu]$; then $H_X(Y)$ can be interpreted as the average of this entropy over x . $H_Y(X)$ has a similar meaning with the roles of the sources $[A, \mu]$ and $[B, \eta]$ interchanged.

In practice, this latter quantity is of the greatest interest. Let us recall that, on the one hand, we regarded the entropy of any finite space as a measure of the uncertainty contained in the space, and, on the other hand, as a measure of the amount of information given by “removing” this uncertainty, i.e., by answering the question of which event of the given space actually occurred. Thus $H_n(X)$ is a measure of the uncertainty in the sequence x_0, \dots, x_{n-1} . The quantity $H_{nY}(X)$ is a measure of the average amount of uncertainty in the same sequence x_0, \dots, x_{n-1} , given that this sequence entered the channel and that the sequence y_0, \dots, y_{n-1} was obtained at the output. Thus $H_{nY}(X)$ indicates how much uncertainty still remains in the sequence x_0, \dots, x_{n-1} after it has been transmitted through the channel. Of course, we always have $H_{nY}(X)=0$ for a noiseless channel, for then knowledge of the received sequence y_0, \dots, y_{n-1} tells us with certainty the transmitted

sequence x_0, \dots, x_{n-1} . In the general case, $H_{nY}(X) > 0$ and represents the “*residual entropy*” of the sequence x_0, \dots, x_{n-1} , which it retains after passing through the channel. After dividing $H_n(X)$ and $H_{nY}(X)$ by n and passing to the limit as $n \rightarrow \infty$, we obtain $H(X)$, the average uncertainty per signal of the source $[A, \mu]$, and $H_Y(X)$, the average uncertainty per signal of the source after the message has passed through the channel.

On the other hand, $H_n(x)$ is the amount of information contained in the sequence x_0, \dots, x_{n-1} , and $H_{nY}(X)$ is the amount of information it retains after being transmitted through the channel. The difference $H_n(X) - H_{nY}(X)$ is thus a measure of the average amount of information given by a sequence x_0, \dots, x_{n-1} transmitted through the channel. After dividing by n and passing to the limit, we find that the difference $H(X) - H_Y(X)$ is the amount of information obtained on the average when a signal of the source $[A, \mu]$ is transmitted through the channel. This quantity

$$R(X, Y) = H(X) - H_Y(X)$$

is therefore the most important characteristic of the quality of transmission. It is natural to call it the *rate* of transmission of information from the given source $[A, \mu]$ through the channel $[A, \nu_x, B]$. We have by (10.3)

$$H_Y(X) = H(X, Y) - H(Y),$$

which means that

$$R(X, Y) = H(X) + H(Y) - H(X, Y).$$

Thus the rate of transmission $R(X, Y)$ is uniquely determined by giving the entropy of the three sources $[A, \mu]$, $[B, \eta]$, and $[C, \omega]$.

#11. The ergodic case

Retaining the terminology and notation of the preceding sections, we now assume that the source $[A, \mu]$ is ergodic and that the channel $[A, \nu_x, B]$ has finite memory m . We shall now show that in this case the compound source $[C, \omega]$ and the source $[B, \eta]$ at the channel output are also ergodic.

Let Z be any cylinder of the space C^t , obtained by fixing any sequence of pairs $(x_0, y_0), \dots, (x_{j-1}, y_{j-1})$ or, equivalently, the pair of sequences

$$u = x_0, \dots, x_{j-1}; \quad v = y_0, \dots, y_{j-1}.$$

The sequences u and v (equivalently, the cylinder Z) will be regarded as constant throughout the proof. Now consider any fixed m -term sequence $z = x_{-m}, \dots, x_{-1}$, and denote by w the sequence $x_{-m} \dots x_{-1} x_0 \dots x_{j-1}$ of constant length $m+j=t$. Since the source $[A, \mu]$ is ergodic, we have almost everywhere in A^t

$$(11.1) \quad \lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=0}^{n-1} g_w(T^k x) = \mu(w).$$

Furthermore, by Birkhoff's theorem, the limit

$$(11.2) \quad \lim_{s \rightarrow \infty} \frac{1}{s} \sum_{t=0}^{s-1} g_w(T^{r+lt} x) = \psi_{r,z}(x)$$

exists almost everywhere in A^t for $0 \leq r \leq t-1$, and for any sequence z . Let us denote by $M \subset A^t$ the set of x for which (11.1) obtains and for which the limits (11.2) exist for any r ($0 \leq r \leq t-1$) and z . Obviously $\mu(M)=1$.

The sum on the left side of (11.2)

$$(11.3) \quad \sum_{t=0}^{s-1} g_w(T^{r+lt} x)$$

is clearly the number of values of l ($0 \leq l \leq s-1$) for which $T^{r+lt}x \in w$, i.e., for which the sequence of length t

$$c_l = x_{r+lt-m}, \dots, x_{r+lt+j-1}$$

coincides with the sequence w . (We note that the sequences c_l fill the sequence $x_{r-m}, \dots, x_{r-m+st-1}$ without gaps and without overlapping, as l runs through the values $0, 1, \dots, s-1$.) Let this number be denoted by

$$p = p(s, r, z, x),$$

where, by (11.2), for any $x \in M$

$$(11.4) \quad p = s\psi_{r,z}(x) + o(s)$$

as $s \rightarrow \infty$, and let l_1, l_2, \dots, l_p , in order of increasing size, denote the values of l ($0 \leq l \leq s-1$) for which c_l coincides with w . Furthermore, let γ_l ($0 \leq l \leq s-1$) denote the (random) sequence

$$y_{r+lt}, \dots, y_{r+lt+j-1}$$

of length j .

The sequences γ_{l_k} ($1 \leq k \leq p$) are of special interest to us. Let A_k denote the event consisting of the sequence γ_{l_k} coinciding with the sequence v (which enters into the definition of the cylinder Z); obviously A_k is the event (cylinder) $T^{r+l_k t}v$ of the space B^t . Its probability for a given x is $\nu_x(A_k) = \nu_x(T^{r+l_k t}v)$. But our channel is without anticipation and has finite memory m . Therefore the event A_k determined by the sequence γ_{l_k} depends statistically only on

$$x_{r+l_k t-m}, \dots, x_{r+l_k t+j-1},$$

i.e., on the sequence c_{l_k} to which x belongs. But, by the definition of the number l_k we have, for $k=1, 2, \dots, p$

$$c_{l_k} = T^{r+l_k t}w.$$

Thus we arrive at the conclusion that the probability $\nu_x(T^{r+l_k t}v)$ has the same value for all $x \in T^{r+l_k t}w$. Since our channel is stationary, this probability is the same as the probability $\nu_x(v)$ of the cylinder v for any $x \in w$. This latter probability, which it is convenient to denote by $\nu_w(v)$, is independent of either k or r .

It follows that $\nu_x(A_k) = \nu_w(v)$ for all k ($1 \leq k \leq p$). Moreover, it is an immediate consequence of the fact that the channel memory is m that the events A_k are independent. We have $p \rightarrow \infty$ as $s \rightarrow \infty$ (if $x \in M$), so that the events A_1, \dots, A_p form a classical Bernouilli scheme. All these events are defined in the space B' . If q denotes the number of values of k ($1 \leq k \leq p$) for which A_k occurs (for a given $x \in M$), then, as $s \rightarrow \infty$ (which, for $x \in M$, implies $p \rightarrow \infty$), we have by the strong law of large numbers

$$(11.5) \quad \frac{q}{p} \rightarrow \nu_w(v)$$

with probability 1. This means that if $x \in M$ and if N denotes the set of values of $y \in B$ for which (11.5) holds as $s \rightarrow \infty$, then $\nu_x(N) = 1$. The set N depends on r , z , and x , i.e., $N = N(r, z, x)$, but r and z can assume only a finite number of different values. If we put

$$\prod_{r,z} N(r, z, x) = N_x^*$$

then obviously for $x \in M$

$$\nu_x(N_x^*) = 1,$$

and (11.5) holds for any r and z , if $x \in M$, $y \in N_x^*$. But since by (11.4)

$$\frac{p}{s} \rightarrow \psi_{r,z}(x) \quad (s \rightarrow \infty)$$

for any $x \in M$, then, for $x \in M$, $y \in N_x^*$, we obtain

$$(11.6) \quad \frac{q}{s} \rightarrow \nu_w(v) \psi_{r,z}(x) \quad (s \rightarrow \infty)$$

for any r and z .

By the definition of the number q , we have immediately

$$q = \sum_{\ell=0}^{s-1} g_w(T^{r+\ell}x) g_v(T^{r+\ell}y)$$

Summing this expression over r from 0 to $t-1$, we find, for $x \in M$, $y \in N_x^*$, using (11.6)

$$\sum_{k=0}^{st-1} g_w(T^kx) g_v(T^ky) = s\nu_w(v) \sum_{r=0}^{t-1} \psi_{r,z}(x) + o(s).$$

But by the definition of the function $\psi_{r,z}(x)$ and by the ergodicity of the source $[A, \mu]$, we have, for $x \in M$

$$\sum_{r=0}^{t-1} \psi_{r,z}(x) = \lim_{s \rightarrow \infty} \frac{1}{s} \sum_{k=0}^{st-1} g_w(T^kx) = t\mu(w),$$

so that

$$(11.7) \quad \sum_{k=0}^{st-1} g_w(T^kx) g_v(T^ky) = st\mu(w) \nu_w(v) + o(s)$$

for $x \in M$, $y \in N_x^*$.

Now we sum (11.7) over all possible sequences z . In doing so, we recall that w is a sequence of the type x_{-m}, \dots, x_{-1} , x_0, \dots, x_{j-1} , where x_{-m}, \dots, x_{-1} is now the variable sequence z , and x_0, \dots, x_{j-1} is the fixed sequence u . Therefore $\sum_z w = u$ and $\sum_z \omega(w, v) = \omega(u, v) = \omega(Z)$. Since $\mu(w)\nu_w(v) = \omega(u, v)$ by the very definition of the distribution ω , the right side of (11.7) becomes

$$st\omega(u, v) + o(s) = st\omega(Z) + o(s)$$

after summation over z . The sum of the terms in the left side of (11.7) can be written

$$\sum_{k=0}^{st-1} g_v(T^k y) \sum_z g_w(T^k x).$$

Since $\sum_z w = u$, the inner sum equals $g_u(T^k x)$; therefore, summing the left side of (11.7) over z gives

$$\sum_{k=0}^{st-1} g_u(T^k x) g_v(T^k y),$$

and we obtain for $x \in M$, $y \in N_x^*$

$$(11.8) \quad \sum_{k=0}^{st-1} g_z(T^k x, T^k y) = st\omega(Z) + o(s).$$

If Q is the set of pairs (x, y) for which $x \in M$, $y \in N_x^*$, then by definition of the distribution ω , and because $\nu_x(N_x^*) = 1$, we have

$$\omega(Q) = \int_M \nu_x(N_x^*) d\mu(x) = \mu(M) = 1.$$

Thus (11.8) obtains almost everywhere in C' . This means that the source $[C, \omega]$ reflects the cylinder Z . Since this cylinder is arbitrary, then by the theorem of #4, we have proved that the source $[C, \omega]$ is ergodic.

It follows at once that the source $[B, \eta]$ is ergodic. Indeed, if $N \in F_B$ is an invariant set in B' , then $A' \times N$ is obviously an invariant set in C' . Since the source $[C, \omega]$ has already been proved to be ergodic, then $\eta(N) = \omega(A' \times N)$ equals 0 or 1; but this means that the source $[B, \eta]$ is ergodic.

CHAPTER IV.

Feinstein's Fundamental Lemma

#12. Formulation of the problem

The authors of most investigations on the foundations of information theory agree in considering the culmination of the theory of discrete information to be Shannon's theorems on the optimum use that can be made of noisy channels by suitably coding the transmitted text. The proofs of these theorems are given only sketchily in Shannon's works, the analysis being limited to sources of the Markov chain type. In his work, McMillan considerably refined in mathematical respects the fundamental concepts of Shannon's theory. (It is just these refined concepts that we used in the preceding sections.) Moreover, he outlined a method of proving Shannon's theorems for any ergodic source; more exactly, he tried to carry over to this case a proof the idea of which had been given by Shannon. In this connection, a number of gaps were discovered, which evidently can not be filled even in the special case of the Markov sources considered by Shannon.

In 1953, Feinstein [4] proposed a fundamentally new way of substantiating Shannon's theorems, which makes the whole theory considerably more transparent. Feinstein's idea consists in deriving from the channel itself as much as can possibly be used to prove Shannon's theorems, before coding and even before connecting the channel to any particular source. This is done by proving a proposition which it is natural to call the fundamental lemma of the whole theory, and which is formulated

without any mention of either sources or codes. The proof of this proposition is our next problem. It should be noted that Feinstein carries out the proof only for channels without memory ($m=0$) and only alludes to possible generalizations. It is, of course, very important to know to how broad a class of channels the fundamental lemma applies. We also note that both the statement and the proof of the lemma as given by Feinstein contain a number of inaccuracies which, however, can be corrected easily.

Consider a stationary channel $[A, \nu_x, B]$, without anticipation and with finite memory m . If this channel is fed by a stationary source $[A, \mu]$, we obtain a definite rate of transmission

$$R(X, Y) = H(X) - H_r(X),$$

which (see #10) can be regarded as the amount of information obtained on the average when one letter is transmitted. Of course, this rate depends on the source A , driving the channel. However, the least upper bound of $R(X, Y)$ for all possible *ergodic* sources $[A, \mu]$ is a quantity which depends only on the channel $[A, \nu_x, B]$ itself. We shall denote this quantity by C , and we shall call it the (*ergodic*) *capacity* of the channel.

As before, we denote the elements of the sets A' and B' by

$$x = \dots, x_{-1}, x_0, x_1, \dots \text{ and } y = \dots, y_{-1}, y_0, y_1, \dots,$$

respectively. Let n be any fixed positive integer. Let us agree to denote by u the sequence (cylinder $x_{-m}, \dots, x_{-1}, x_0, \dots, x_{n-1}$) of length $n+m$, where all the x_i take on definite values (letters of the alphabet A). Obviously, the number of different sequences u is a^{m+n} , where a is the number of letters in the alphabet A . Similarly, let us denote by v the sequence (cylinder) y_0, \dots, y_{n-1} of length n , where all the y_i

are letters of the alphabet B . Obviously, the number of different sequences v is b^n , where b is the number of letters in the alphabet B . Since the memory of the channel is m , then the probability $\nu_x(v)$ of receiving $y \in v$ at the channel output when a given x is transmitted is the same for all x belonging to the same sequence u . Thus, this probability depends only on the sequences u and v selected, and it is natural to denote it by $\nu_u(v)$. Similarly, if V is the union of several sequences v , we shall denote by $\nu_u(V)$ the probability of the event $y \in V$ when any $x \in u$ is transmitted.

Now, let λ be any constant $(0 < \lambda < \frac{1}{2})$. Let us agree to call a group $\{u_i\}$ ($1 \leq i \leq N$) of u -sequences *distinguishable* if there exists a group $\{V_i\}$ ($1 \leq i \leq N$) of sets $V_i \subset B^l$ where each V_i ($1 \leq i \leq N$) is a set of several sequences v , such that 1) V_i and V_k have no sequences in common for $i \neq k$, and 2) $\nu_{u_i}(V_i) > 1 - \lambda$ ($1 \leq i \leq N$). Clearly, this definition of a distinguishable group depends on the parameter λ . We can now formulate Feinstein's fundamental lemma as follows.

Lemma.

If a given channel is stationary, without anticipation, and with finite memory m , then, for sufficiently small $\lambda > 0$ and sufficiently large n , there exists a distinguishable group $\{u_i\}$ ($1 \leq i \leq N$) of u -sequences with

$$N > 2^{n(C-\lambda)}$$

members, where C is the (ergodic) capacity of the channel.

The importance of a proposition of this kind in evaluating the optimum transmission of information is clear almost immediately. If λ is small, since $\nu_{u_i}(V_i) > 1 - \lambda$, when a sequence u_i is transmitted, then at the channel output we obtain with

overwhelming probability a sequence v in the group V_i ; since the groups V_i have no v in common, then, knowing v , we uniquely determine the group V_i which contains it, which means that we guess with an overwhelming probability the transmitted sequence u_i . Of course all this is under the condition that the group $\{u_i\}$ is distinguishable. Thus, if only sequences of a certain distinguishable group are used to send signals, then, despite the noise, the signals sent can be guessed with overwhelming probability. The question of whether any text suitable for transmission can be "encoded" into the sequences u_i depends, of course, on how many such sequences there are in the first place. The fundamental lemma shows at once that the number of such sequences is very great for sufficiently large n . After this lemma has been proved, we shall consider the question of how to use the estimate of N that it affords to compute an index of optimum transmission of given material through a given channel. In fact, this is the way we shall approach the basic Shannon theorems.

#13. Proof of the lemma

We shall divide the proof into a number of separate steps in the interest of a better presentation.

1. *The source $[A, \mu]$.* By definition of the number C , we can find an ergodic source $[A, \mu]$ for the channel $[A, \nu_x, B]$ such that when this source drives the channel, we have a rate of transmission (see the end of #10)

$$R(X, Y) = H(X) - H_r(X) > C - \frac{\lambda}{4}.$$

Here $H(X)$ denotes the entropy of the source $[A, \mu]$ and $H_r(X)$ the average conditional entropy of the same source, given the

signal received at the channel output. The source $[A, \mu]$, being ergodic, has the E property according to McMillan's theorem (see #5). This means the following. Let w denote an arbitrary sequence x_0, \dots, x_{n-1} of letters from the alphabet A (w is therefore some cylinder in A^I). Moreover, let $W_0 \subset A^I$ denote the set (union) of all cylinders w for which

$$(13.1) \quad \left| \frac{\lg \mu(w)}{n} + H(X) \right| \leq \frac{\lambda}{4},$$

then, for arbitrarily small $\lambda > 0$ we have for sufficiently large n

$$\mu(W_0) > 1 - \lambda.$$

2. *The source $[C, \omega]$.* Connecting the source $[A, \mu]$ to the channel $[A, \nu_x, B]$ gives us according to the definitions of #10 the *compound* source $[C, \omega]$, with the alphabet $C = A \times B$ and entropy $H(X, Y)$. By the theorems of #11, this source is ergodic, and therefore, by McMillan's theorem (#5) has the E property. This means the following. Each pair (w, v) of n -term sequences $[w \subset A^I, v \subset B^I]$ is some cylinder in the space C^I , with a definite probability $\omega(w, v)$ in this space. Let Z denote the set of all cylinders (w, v) for which

$$(13.2) \quad \left| \frac{\lg \omega(w, v)}{n} + H(X, Y) \right| \leq \frac{\lambda}{4};$$

then, for arbitrarily small $\lambda > 0$ and $\delta > 0$, we have, for sufficiently large n

$$\omega(Z) > 1 - \delta.$$

3. *The source $[B, \eta]$.* This source, which we described in #10, is determined by the relation

$$\eta(M) = \int_{A^I} \nu_x(M) d\mu(x) = \omega(A^I \times M) \quad (M \in F_B).$$

We proved that it is ergodic at the end of #11. According to McMillan's theorem it has the E property. This means the following. The probability of the set of n -term sequences $v \subset B'$ for which

$$(13.3) \quad \left| \frac{\lg \eta(v)}{n} + H(Y) \right| \leq \frac{\lambda}{4}$$

is as close to 1 as desired for sufficiently large n . ($H(Y)$ denotes the entropy of the source $[B, \eta]$.)

4. *The set A_w and its probability.* Let us denote by X the set of pairs of sequences (w, v) for which the inequalities (13.2) and (13.3) are satisfied. By the foregoing, we have

$$\omega(X) > 1 - \frac{1}{2} \lambda^2,$$

if n is large enough. Now, for any sequence w , let $A_w \subset B'$ denote the set of all sequences v for which $(w, v) \in X$, i.e., for which the inequalities (13.2) and (13.3) are satisfied. Furthermore, let $W_1 \subset A'$ denote the set of all sequences $w \subset W_0$ for which

$$(13.4) \quad \frac{\omega(w, A_w)}{\mu(w)} = \frac{\omega(w, A_w)}{\omega(w, B')} > 1 - \frac{\lambda}{2}.$$

We now estimate the probability of the set W_1 by using Lemma 2.1. The roles of the spaces A , B , AB are played by the sets of all sequences w , v , (w, v) , respectively. For Z we take the set X , for U_0 the set W_0 . Then clearly $\frac{1}{2} \lambda^2$ plays the role of δ_1 and λ that of δ_2 . The set of sequences v for which $(w, v) \notin X$, i.e., $\Gamma_w = B' - A_w$, takes the place of the set Γ_w . Finally, the set of $w \subset W_0$ for which the conditional probability

$$P_w(\Gamma_w) = 1 - P_w(A_w) = 1 - \frac{\omega(w, A_w)}{\mu(w)} \leq \alpha$$

plays the role of the set U_1 . Putting $\alpha = \frac{\lambda}{2}$, we can take our set W_1 as the set U_1 of Lemma 2.1. We see at once that all the premises of Lemma 2.1 have been fulfilled. Using it, we conclude that

$$(13.5) \quad \mu(W_1) > 1 - \lambda - \lambda = 1 - 2\lambda.$$

Now let the sequence w belong to the set W_1 and the sequence v to the set A_w , so that by the definition of the set A_w we have $(w, v) \subset X$. Since $W_1 \subset W_0$, then, by the definition of the sets X and W_0 , all three inequalities (13.1), (13.2), (13.3) are fulfilled for the pairs (w, v) . In particular, we have

$$\begin{aligned} \frac{\lg \mu(w)}{n} + H(X) &\leq \frac{\lambda}{4}; & \frac{\lg \eta(v)}{n} + H(Y) &\leq \frac{\lambda}{4}; \\ \frac{\lg \omega(w, v)}{n} + H(X, Y) &\geq -\frac{\lambda}{4}, \end{aligned}$$

whence

$$\lg \frac{\omega(w, v)}{\mu(w) \eta(v)} + n[H(X, Y) - H(X) - H(Y)] \geq -\frac{3}{4}n\lambda.$$

Since (see #10)

$$H(X) + H(Y) - H(X, Y) = R(X, Y),$$

we find

$$\lg \frac{\omega(w, v)}{\mu(w) \eta(v)} \geq n \left[R(X, Y) - \frac{3}{4}\lambda \right],$$

and therefore

$$\frac{\omega(w, v)}{\mu(w)} \geq 2^{n[R(X, Y) - \frac{3}{4}\lambda]} \eta(v).$$

Since we have $R(X, Y) > C - \frac{\lambda}{4}$ by the choice of the source $[A, \mu]$, then

$$\frac{\omega(w, v)}{\mu(w)} > 2^{n(C-\lambda)} \eta(v).$$

This inequality holds for any $w \subset W_1$, $v \subset A_w$. Keeping the sequence $w \subset W_1$ fixed, and summing over all $v \subset A_w$, we find

$$\frac{\omega(w, A_w)}{\mu(w)} > 2^{n(C-\lambda)} \eta(A_w),$$

but the left side of this inequality does not exceed unity, since $\omega(w, A_w) \leq \omega(w, B^r) = \mu(w)$. Therefore we find that for any sequence $w \subset W_1$,

$$(13.6) \quad \eta(A_w) < 2^{-n(C-\lambda)}.$$

5. *Special groups of w-sequences.* We agree to call the group $\{w_i\} (1 \leq i \leq N)$ of w -sequence *special* if a set B_i of v -sequences can be associated with each sequence w_i of this group such that

- 1) the sets B_i and B_k have no common sequence for $i \neq k$;
- 2) $\frac{\omega(w_i, B_i)}{\mu(w_i)} > 1 - \lambda \quad (1 \leq i \leq N);$
- 3) $\eta(B_i) < 2^{-n(C-\lambda)} \quad (1 \leq i \leq N).$

First of all, let us convince ourselves that *special* groups exist. To do this, we take any sequence $w \subset W_1$ and put $B = A_w$. Then by (13.4) and (13.6), we have

$$\frac{\omega(w, B)}{\mu(w)} > 1 - \lambda; \quad \eta(B) \leq 2^{-n(C-\lambda)}.$$

Therefore, every sequence $w \subset W_1$ is a special group, which proves the existence of special groups. Now let us call a

special group of w -sequences *maximal*, if adding any other sequence to it causes it to lose the character of a special group.* In view of the foregoing, the existence of a maximal group is obvious.

6. *Estimate of the number of members of a maximal special group.* Let $M = \{w_i\}$ ($1 \leq i \leq N$) be a maximal special group of w -sequences. Now, for any sequence w , let us put

$$A_w - A_w \sum_{i=1}^N B_i = B_w.$$

Obviously the set $B_w \subset B^I$ has no elements in common with any of the sets B_i ($1 \leq i \leq N$). If $w \subset W_1$, then by (13.6)

$$\eta(B_w) \leq \eta(A_w) < 2^{-n(C-\lambda)}.$$

If the sequence $w \subset W_1$ did not belong to the group M and if we had at the same time

$$\frac{\omega(w, B_w)}{\mu(w)} > 1 - \lambda,$$

then adding this sequence w to the group M would evidently again give a special group of w -sequences, which is impossible, since the group M is assumed to be maximal. Therefore, if $w \subset W_1$ and $w \notin M$ (i.e., if $w \subset W_1 - MW_1$), then

$$(13.7) \quad \frac{\omega(w, B_w)}{\mu(w)} \leq 1 - \lambda.$$

But by the definition of the set B_w

$$\omega(w, B_w) = \omega(w, A_w) - \omega\left(w, A_w \sum_{i=1}^N B_i\right),$$

and if $w \subset W_1 - MW_1$, it follows from (13.4) and (13.7) that

* If the set of all w -sequences is a special group, then we consider this group to be maximal also.

$$\begin{aligned}\omega\left(w, A_w \sum_{i=1}^N B_i\right) &= \omega(w, A_w) - \omega(w, B_w) \\ &> \left(1 - \frac{\lambda}{2}\right) \mu(w) - (1 - \lambda) \mu(w) = \frac{\lambda}{2} \mu(w).\end{aligned}$$

From this we find that

$$\begin{aligned}\eta\left(\sum_{i=1}^N B_i\right) &= \omega\left(A^I, \sum_{i=1}^N B_i\right) \geq \omega\left(W_1, \sum_{i=1}^N B_i\right) \\ &= \omega\left(W_1 - MW_1, \sum_{i=1}^N B_i\right) + \omega\left(MW_1, \sum_{i=1}^N B_i\right) \\ (13.8) \quad &= \sum_{w \subset W_1 - MW_1} \omega\left(w, \sum_{i=1}^N B_i\right) + \sum_{w \subset MW_1} \omega\left(w, \sum_{i=1}^N B_i\right) \\ &> \frac{\lambda}{2} \sum_{w \subset W_1 - MW_1} \mu(w) + \sum_{w \subset MW_1} \omega\left(w, \sum_{i=1}^N B_i\right) \\ &\geq \frac{\lambda}{2} [\mu(W_1) - \mu(MW_1)] + (1 - \lambda) \mu(MW_1),\end{aligned}$$

where only the estimate of the last sum requires explanation. The fact is that if $w \subset MW_1 \subset M$, the sequence w is one of the sequences w_i of the special group M . For example, if $w = w_k$ then

$$\omega\left(w, \sum_{i=1}^N B_i\right) \geq \omega(w_k, B_k) > (1 - \lambda) \mu(w),$$

and therefore

$$\sum_{w \subset MW_1} \omega\left(w, \sum_{i=1}^N B_i\right) > (1 - \lambda) \mu(MW_1).$$

Furthermore, using (13.5) we find that for $\lambda \leq \frac{1}{2}$, (13.8) becomes

$$\begin{aligned}(13.9) \quad \eta\left(\sum_{i=1}^N B_i\right) &\geq \frac{\lambda}{2} \mu(W_1) + \left(1 - \lambda - \frac{\lambda}{2}\right) \mu(MW_1) \\ &\geq \frac{\lambda}{2} \mu(W_1) > \frac{\lambda}{2} (1 - 2\lambda) = \gamma.\end{aligned}$$

But since by the definition of a *special* group

$$\eta(B_i) < 2^{-n(C-\lambda)} \quad (1 \leq i \leq N),$$

then, on the other hand

$$\eta \left(\sum_{i=1}^N B_i \right) < N \cdot 2^{-n(C-\lambda)}.$$

Comparing this with the inequality (13.9) we find

$$N > \gamma 2^{n(C-\lambda)},$$

and therefore, for sufficiently large n

$$N > 2^{n(C-2\lambda)}.$$

Thus we have established an important lower bound for the number of terms in any maximal special group of w -sequences.

7. *Completion of the proof of the fundamental lemma.* Let us consider an arbitrary maximal special group $\{w_i\}$ ($1 \leq i \leq N$) of w -sequences, and let us extend each sequence w_i of this group by adding m more letters to the left, obtaining in this way a new sequence u_i of length $n+m$. Let us convince ourselves that by selecting this extension in a suitable way, we have

$$\frac{\omega(u_i, B_i)}{\mu(u_i)} > 1 - \lambda \quad (1 \leq i \leq N).$$

Every possible extension of the sequence w_i by m letters to the left gives a cylinder (sequence) u_i contained in the cylinder w_i ($u_i \subset w_i$), and the union of all the cylinders $u_i \subset w_i$ is the cylinder w_i . Therefore for $1 \leq i \leq N$

$$\frac{\omega(w_i, B_i)}{\mu(w_i)} = \sum_{u_i \subset w_i} \frac{\omega(u_i, B_i)}{\mu(w_i)} = \sum_{u_i \subset w_i} \frac{\omega(u_i, B_i)}{\mu(u_i)} \frac{\mu(u_i)}{\mu(w_i)}.$$

The left side of this equality exceeds $1 - \lambda$ by the definition of a special group. The sum of the second factors in the sum on the right side is obviously unity. Therefore at least one of

the first factors must exceed $1-\lambda$; but this means that, for at least one extension, we must have

$$(13.10) \quad \frac{w(u_i, B_i)}{\mu(u_i)} > 1 - \lambda,$$

as asserted.

By the definition of the distribution ω

$$\omega(u_i, B_i) = \int_{u_i} \nu_x(B_i) d\mu(x).$$

But, as we saw in #12, $\nu_x(B_i)$ assumes the same value for all $x \in u_i$, which we agreed to denote by $\nu_{u_i}(B_i)$. Thus

$$\omega(u_i, B_i) = \mu(u_i) \nu_{u_i}(B_i).$$

and (13.10) can be written as

$$\nu_{u_i}(B_i) > 1 - \lambda \quad (1 \leq i \leq N).$$

Selecting the extension of the sequence w_i in the manner indicated above, we obtain a group of N sequences u_i of length $n+m$ (of letters from the alphabet A) and a group of N sets B_i , each of which is the union of several sequences v of length n (of letters from the alphabet B), where 1) the sets B_i and B_k have no common elements for $i \neq k$, 2) $\nu_{u_i}(B_i) > 1 - \lambda$ ($1 \leq i \leq N$), and 3) $N > 2^{n(C-2\lambda)}$. This just means that the sequences u_i ($1 \leq i \leq N$) form a distinguishable group. But since the number of terms of this group exceeds $2^{n(C-2\lambda)}$ and $\lambda > 0$ can be chosen as small as desired, the proof of the fundamental lemma is complete.

CHAPTER V.

Shannon's Theorems

#14. Coding

Up to now, we have always assumed that the source feeding the given channel has an alphabet which is identical with the input alphabet of the channel. However, in practice these alphabets are different in the majority of cases, and we must now consider the general case.

Suppose the output of the stationary source $[A_0, \mu]$ is to be transmitted by means of the stationary channel $[A, \nu_x, B]$ where in general the alphabets A_0 and A are different. Then, before transmitting, it is necessary to transform ("encode") the sequence of letters from the alphabet A_0 emanating from the source into some sequence of letters from the alphabet A . As usual, we suppose that any message emanating from the source $[A_0, \mu]$ is in the form of a sequence

$$(14.1) \quad \theta = \dots, \theta_{-1}, \theta_0, \theta_1, \dots,$$

where all the θ_i are letters of the alphabet A_0 . We uniquely transform ("encode") each θ into some sequence

$$(14.2) \quad x = \dots, x_{-1}, x_0, x_1, \dots,$$

where all the x_i are letters of the alphabet A . The rules of this transformation constitute the "code" used. Thus, from the mathematical point of view, a code is simply any function

$$x = x(\theta),$$

where $\theta \in A_0^I$, $x \in A^I$. It is clear that formally every code can

be regarded as a kind of channel with input alphabet A_0 and output alphabet A . Such channels are characterized as being noiseless, i.e., to a sequence θ at the channel input corresponds a unique sequence $x=x(\theta)$ at the output. We can give this channel the usual designation $[A_0, \rho_0, A]$, where the form of the function $\rho_0(M)$ ($M \in A^I$) is also apparent, namely

$$\rho_0(M) = \begin{cases} 1 & [x(\theta) \in M] \\ 0 & [x(\theta) \notin M] \end{cases}.$$

However, not every code has practical value. If we must actually know (as will be necessary in the general case) the whole sequence θ , i.e., the whole infinite message from the source $[A_0, \mu]$ in order to determine any letter x_k of the encoded text, then obviously, in practice, we can never determine this letter. Consequently, the only codes of importance in applications are those such that it suffices to know some finite sequence of letters $\theta_k \in A_0$ in order to uniquely determine each letter $x_k \in A$ (and also each finite sequence of such letters). In particular, in information theory “sequential coding”, is used predominantly, which consists of the following: Both the sequences (14.1) and (14.2) are divided into finite subsequences of any length which are numbered from left to right, just like the letters of these sequences. A rule is given which uniquely determines the k 'th subsequence of the sequence x in terms of the k 'th subsequence of the sequence θ .

Now let us turn to the problem of transmitting the output of the source $[A_0, \mu]$ by means of the channel $[A, \nu_x, B]$. We code each message θ of the source $[A_0, \mu]$ into some specific message x , composed of letters from the alphabet A . We then pass this x through the channel $[A, \nu_x, B]$ and obtain some

message $y \in B'$ at its output. Obviously, the result of connecting the code selected to the channel $[A, \nu_x, B]$ can be regarded as a new channel $[A_0, \lambda_0, B]$. It is very easy to determine the function (probability) $\lambda_0(Q)$ ($Q \in F_B$). Since the coding transforms the message $\theta \in A_0^I$ into the message $x(\theta) \in A'$, the probability $\lambda_0(Q)$ of obtaining $y \in Q$ for a given θ is the probability of obtaining $y \in Q$ if the message $x=x(\theta)$ enters the input of the channel $[A, \nu_x, B]$, i.e.

$$\lambda_0(Q) = \nu_{x(\theta)}(Q),$$

and the channel $[A_0, \lambda_0, B]$ can be written as $[A_0, \nu_{x(\theta)}, B]$.

The process which we are examining consists in connecting the source $[A_0, \mu]$ to this new channel $[A_0, \lambda_0, B]$; this time the source alphabet coincides with the channel input alphabet, so that we have a situation with which we are familiar. In particular, connecting the source to the channel gives, as we know, the “compound” source $[C, \omega]$ where $C = A \times B$, and where the probability distribution is such that for $M \in F_{A_0}$, $N \in F_B$ we have

$$\omega(M \times N) = \int_M \lambda_0(N) d\mu(\theta) = \int_M \nu_{x(\theta)}(N) d\mu(\theta).$$

#15. The first Shannon theorem

Suppose we have an ergodic source $[A_0, \mu]$ and a stationary, non-anticipating channel $[A, \nu_x, B]$ with finite memory m . Let the entropy of the source $[A_0, \mu]$ be H_0 , and let the ergodic capacity of the channel $[A, \nu_x, B]$ be C . We assume that $H_0 < C$, and choose a positive number $\lambda < \frac{1}{2}(C - H_0)$. Then we select a

particular code which transforms the output of the source $[A_0, \mu]$ into the alphabet A of the given channel.

First of all, we note that the source $[A_0, \mu]$, being ergodic, has the E property (see §5). In particular, this means that for sufficiently large n , the sequences α of letters from the alphabet A_0 can be divided into two groups, a “high probability” group, for each sequence α of which

$$\frac{\lg \mu(\alpha)}{n} + H_0 > -\lambda,$$

or, equivalently

$$\mu(\alpha) > 2^{-n(H_0 + \lambda)},$$

and a “low probability” group, the total probability of which is as small as desired. Obviously the number of sequences in the high probability group is less than $2^{n(H_0 + \lambda)} < 2^{n(C - \lambda)}$. In what follows we shall denote these sequences by $\alpha_1, \alpha_2, \dots$. We shall denote the set of all sequences in the low probability group by α_0 .

Now we turn to the channel $[A, \nu_x, B]$. By the fundamental lemma of Chapter IV, for sufficiently large n there exists a distinguishable group $\{u_i\}$ ($1 \leq i \leq N$) of sequences u of length $n+m$ (consisting of letters from the alphabet A), with $N > 2^{n(C - \lambda)}$ members. Thus, since N is larger than the number of sequences α_i in the high probability group just considered, we can associate with each sequence α_i a sequence u_i such that different u_i correspond to different α_i . In doing this, at least one sequence of the distinguishable group has not been used; we associate this sequence with all the sequences α_0 of the low probability group. After doing this, to each n -term sequence α of letters from the alphabet A_0 corresponds a specific sequence u of

length $n+m$ from the alphabet A , a sequence which belongs to the distinguishable group $\{u_i\}$.

We now divide the sequence θ of letters from the alphabet A_0 , which is to be coded, into subsequences of length n , and the sequence x , into which θ is to be encoded, into subsequences of length $n+m$. We number both sets of subsequences from left to right, as usual. The k 'th subsequence in the message θ will be one of the sequences α ; then we select as the k 'th subsequence in the message x , the sequence u_i of our distinguishable group which corresponds to this sequence α . Doing this for all k ($-\infty < k < +\infty$), it is clear that we uniquely determine $x=x(\theta)$ in terms of the given θ . In this way we have set up a specific code, which we shall retain in all that follows. Obviously, this code is an example of the "sequential coding" discussed in #14.

Connecting the code just selected to the channel $[A, \nu_x, B]$ we obtain, as we saw in #14, a new channel $[A_0, \lambda_0, B]$, where we have $\lambda_0(Q)=\nu_{x(\theta)}(Q)$ for $Q \in F_B$. Making the source $[A_0, \mu]$ feed this new channel, we arrive, according to #14, at the "compound" source $[C, \omega]$, where $C=A_0 \times B$ and ω ($M \times N$) (where $M \in F_{A_0}$, $N \in F_B$) is given by the last formula of #14. Let us agree to denote the different sequences of length n , consisting of letters from the alphabet B , by β_k ($k=1, 2, \dots$). Then, by the definition of the distribution ω , we have for any $k \geq 1$ and $i=0, 1, \dots, N$

$$\omega(\alpha_i \times \beta_k) = \int_{\alpha_i} \lambda_0(\beta_k) d\mu(\theta) = \int_{\alpha_i} \nu_{x(\theta)}(\beta_k) d\mu(\theta).$$

But $\theta \in \alpha_i$ is equivalent to $x(\theta) \in u_i^*$, and since $\nu_x(\beta_k)$ assumes

* Here u_0 is understood to be the sequence of the distinguishable group $\{u_i\}$ into which the whole "low probability group" α_0 is encoded.

the same value for all $x \in u_i$, which we agreed in #12 to designate by $\nu_{u_i}(\beta_k)$, then we have $\nu_{x(0)}(\beta_k) = \nu_{u_i}(\beta_k)$ for all $\theta \in \alpha_i$. Thus we find

$$(15.1) \quad \omega(\alpha_i \times \beta_k) = \mu(\alpha_i) \nu_{u_i}(\beta_k).$$

In order to avoid any misunderstanding, we recall the meaning of the event $\alpha_i \times \beta_k$, the probability of which is given by (15.1): $\alpha_i \times \beta_k$ denotes the joint event that 1) the n -term sequence transmitted by the source $[A_0, \mu]$ coincides with α_i (if $i > 0$) or belongs to the low probability group (if $i = 0$), and 2) after being transmitted through the channel $[A_0, \lambda_0, B]$, the n -term sequence transmitted by the source $[A_0, \mu]$ gives an $n+m$ -term sequence of letters from the alphabet B , the last n letters of which are the sequence β_k .

Now let i_k be the value of the subscript i ($0 \leq i \leq N$) for which the probability $\omega(\alpha_i, \beta_k)^*$ has its greatest value. (If there are several such values i , then we take any of them as i_k). Since the conditional probability of the sequence α_i for a given sequence β_k equals

$$\frac{\omega(\alpha_i, \beta_k)}{\omega(A_0^I, \beta_k)},$$

where the denominator is independent of i , then we can also say α_{i_k} is the sequence α_i which is most probable for a given sequence β_k . Let us set

$$\sum_k \sum_{i \neq i_k} \omega(\alpha_i, B_k) = P.$$

Obviously, we can regard P as the probability that the sequence α_i at the input of the channel $[A_0, \lambda_0, B]$ will not be the most probable one for a given sequence β_k at the output of the channel.

* $\omega(\alpha_i, \beta_k)$ and $\omega(\alpha_i \times \beta_k)$ have the same meaning.

The sequences u_i , into which we encoded all the sequences α_i form a distinguishable group; this implies the existence of a group $\{B_i\}$ ($1 \leq i \leq n$) (where every B_i is the union of several sequences β_k) such that 1) $\nu_{u_i}(B_i) > 1 - \lambda$, and 2) B_i and B_j have no sequences in common for $i \neq j$. But summing (15.1) over all $\beta_k \subset B_i$, we find

$$\omega(\alpha_i, B_i) = \mu(\alpha_i) \nu_{u_i}(B_i) > (1 - \lambda) \mu(\alpha_i) \quad (0 \leq i \leq N).$$

Therefore the conditional probability of obtaining a sequence $\beta_k \subset B_i$ at the output of the channel $[A_0, \lambda_0, B]$, given that we had a sequence α_i at the input of the channel is

$$P_{\alpha_i}(B_i) = \frac{\omega(\alpha_i, B_i)}{\mu(\alpha_i)} > 1 - \lambda \quad (0 \leq i \leq N).$$

Thus we see that the two complete systems of events α_i ($0 \leq i \leq N$) and β_k ($k = 1, 2, \dots$) satisfy all the premises of Lemma 2.2. Using this lemma, we find

$$P \leq \lambda.$$

This important result is the first Shannon theorem, which we can state as follows.

Theorem.

Let there be given 1) a stationary, non-anticipating channel $[A, \nu_x, B]$ with ergodic capacity C and finite memory m , and 2) an ergodic source $[A_0, \mu]$ with entropy $H_0 < C$. Let $\varepsilon > 0$. Then, for sufficiently large n , the output of the source $[A_0, \mu]$ can be encoded into the alphabet A in such a way that each sequence α_i of n letters from the alphabet A_0 is mapped into a sequence u_i of $n+m$ letters from the alphabet A , and such that if the sequence u_i is transmitted through the given channel, we can determine the transmitted sequence α_i with a probability

greater than $1-\varepsilon$ from the sequence received at the channel output.

The definition of the sequence α_i entails that we select as α_i the sequence which is most probable, given the n letters last received at the channel output. The first Shannon theorem is often formulated less exactly by saying that for $H_0 < C$ it is always possible to find a code such that the transmitted sequence can be guessed from the received sequence with a probable error which does not exceed an arbitrarily small number ε .

#16. The second Shannon theorem

We retain, *in toto*, the source $[A_0, \mu]$, the channel $[A, \nu_x, B]$ and the code of the preceding section, and also all the notation introduced there. However, we now consider a new problem pertinent to this whole process, namely, the evaluation of its transmission rate, i.e., the amount of information given on the average by one letter at the channel output. To this end, let us examine first the finite probability space $\alpha_i \beta_k$ ($0 \leq i \leq N$, $k=1,2,\dots$), with the distribution $\omega(\alpha_i, \beta_k)$, which was considered in #15. We recall that α_i was one of the n -term sequences of the high probability group in the output of the given source for $i \geq 1$, that α_0 was the whole low probability group of such sequences and that β_k was an n -term sequence of letters from the alphabet B (at the channel output). We denoted by P the quantity

$$P = \sum_k \sum_{i \neq i_k} \omega(\alpha_i, \beta_k),$$

where i_k is the value of the subscript i for which the probability $\omega(\alpha_i, \beta_k)$ is the greatest, and we proved that $P < \lambda$.

It is clear that we find ourselves in a position to apply Lemma

2.3; the spaces $\{\alpha_i\}$ and $\{\beta_k\}$ play the role of the spaces $\{A_i\}$ and $\{B_k\}$ of this lemma, and the number $N+1$ plays the role of the number n . If, as usual, we denote by $H_p(\alpha)$ the conditional entropy of the space $\{\alpha_i\}$ for a given β_k , averaged over all β_k , then by Lemma 2.3 we have

$$H_p(\alpha) \leq P \lg N - P \lg P - (1-P) \lg (1-P).$$

Here the number N is the number of sequences in the high probability group; as we saw in #15, it does not exceed $2^{n(C-\lambda)} < 2^{nC}$, whence $\lg N < nC$. Taking into account that $P < \lambda$ and that, as is easily seen $-P \lg P - (1-P) \lg (1-P) < 1$ for any P ($0 < P < 1$), we obtain

$$H_p(\alpha) < \lambda nC + 1.$$

In particular, since λ can be chosen arbitrarily small for sufficiently large n , we obtain

$$H_p(\alpha) = o(n) \quad (n \rightarrow \infty).$$

Here

$$H_p(\alpha) = - \sum_k \eta(\beta_k) \sum_{i=0}^N f[p_{\beta_k}(\alpha_i)],$$

where

$$f(x) = x \lg x; \quad \eta(\beta_k) = \omega(A^I, \beta_k) \quad \text{and} \quad p_{\beta_k}(\alpha_i) = \frac{\omega(\alpha_i, \beta_k)}{\omega(A^I, \beta_k)} = \frac{\omega(\alpha_i, \beta_k)}{\eta(\beta_k)}.$$

In all the foregoing α_0 , unlike α_i for $i > 0$, denotes not an individual sequence, but the whole “low probability” group of n -term sequences made up of letters from the alphabet A . Now we divide this group into separate sequences

$$\alpha'_1, \alpha'_2, \dots, \alpha'_q$$

and instead of the space (α_i) ($0 \leq i \leq N$) we consider the space

(α_i, α'_j) ($1 \leq i \leq N, 1 \leq j \leq q$) of all n -term sequences consisting of letters from the alphabet A_0 , with appropriate probabilities $\mu(\alpha_i), \mu(\alpha'_j)$. We leave the space (β_k) as before. Denoting by $H_{\beta}(\alpha, \alpha')$ the conditional entropy of the space (α_i, α'_j) for a given β_k , averaged over β_k , we find

$$(16.1) \quad \begin{aligned} H_{\beta}(\alpha, \alpha') &= - \sum_k \eta(\beta_k) \left\{ \sum_{i=1}^N f[p_{\beta_k}(\alpha_i)] + \sum_{j=1}^q f[p_{\beta_k}(\alpha'_j)] \right\} \\ &= H_{\beta}(\alpha) + \sum_k \eta(\beta_k) f[p_{\beta_k}(\alpha_0)] + R, \end{aligned}$$

where

$$(16.2) \quad R = - \sum_k \eta(\beta_k) \sum_{j=1}^q f[p_{\beta_k}(\alpha'_j)].$$

(In the interest of complete clarity, we note that here

$$p_{\beta_k}(\alpha'_j) = \frac{\omega(\alpha'_j, \beta_k)}{\eta(\beta_k)},$$

where

$$\omega(\alpha'_j, \beta_k) = \int_{\alpha'_j} \lambda_0(\beta_k) d\mu(\theta).$$

Since $\lambda_0(\beta_k)$ has the same value $\lambda_{\alpha_0}(\beta_k)$ for all $\theta \in \alpha'_j \subset \alpha_0$, then

$$p_{\beta_k}(\alpha'_j) = \frac{\lambda_{\alpha_0}(\beta_k) \mu(\alpha'_j)}{\eta(\beta_k)}$$

is uniquely given by the data of our problem.)

Since the second term in the right side of (16.1) is negative, then

$$(16.3) \quad H_{\beta}(\alpha, \alpha') < H_{\beta}(\alpha) + R,$$

and since we have already estimated $H_{\beta}(\alpha)$, it remains only to estimate the quantity R defined by the sum (16.2). Using Lemma 1.1, we find

$$(16.4) \quad R \leq - \sum_{j=1}^q \mu(\alpha'_j) \lg \mu(\alpha'_j),$$

Here, by definition of the low-probability group, we have $\sum_{j=1}^q \mu(\alpha'_j) = \mu(\alpha_0) < \lambda$. But it is easily seen that the greatest value of the sum (16.4), under the supplementary condition

$$\sum_{j=1}^q \mu(\alpha'_j) = \varepsilon,$$

is attained for $\mu(\alpha'_j) = \frac{\varepsilon}{q}$ ($1 \leq j \leq q$) and is $\varepsilon \left[\lg q + \lg \frac{1}{\varepsilon} \right]$.

Therefore, in our case

$$R < \lambda \left[\lg q + \lg \frac{1}{\varepsilon} \right].$$

But q is the number of sequences in the low probability group and is less than a^n , the number of all n -term sequences, (a is the number of letters in the alphabet A_0 .) Therefore

$$R < \lambda n \lg a + \lambda \lg \frac{1}{\varepsilon} < \lambda n \lg a + 1.$$

We see that $R = o(n)$ as $n \rightarrow \infty$, and since $H_\beta(\alpha) = o(n)$, as we proved earlier, it follows from (16.3) that

$$(16.5) \quad H_\beta(\alpha, \alpha') = o(n) \quad (n \rightarrow \infty).$$

We now change our notation somewhat. By the space $\alpha(\alpha_1, \alpha_2, \dots)$ we shall mean the set of all n -term sequences of letters from the alphabet A_0 , whether from the high probability or low probability groups, so that the number of all α_i is a^n , where a is the number of letters in the alphabet A_0 . As before, the space $\beta(\beta_1, \beta_2, \dots)$ denotes the set of all n -term sequences of letters from the alphabet B . The product (α, β) of these two spaces has the distribution

$$\omega(\alpha_i, \beta_k) = \int_{\alpha_i} \lambda_\theta(\beta_k) d\mu(\theta) = \lambda_{\alpha_i}(\beta_k) \mu(\alpha_i),$$

and the distributions of the spaces α and β are given by the functions $\mu(\alpha_i)$ and

$$\eta(\beta_k) = \omega(A_0^t, \beta_k) = \sum_i \mu(\alpha_i) \lambda_{\alpha_i}(\beta_k),$$

respectively. The conditional entropy of the space α for a given β_k , averaged over β_k , which we denote by $H_\beta(\alpha)$ is obviously the quantity $H_\beta(\alpha, \alpha')$ which we have just estimated. Therefore, by (16.5) we have

$$(16.6) \quad H_\beta(\alpha) = o(n) \quad (n \rightarrow \infty).$$

The information transmission process which we are considering is the following. The output of the source $[A_0, \mu]$ is “cut up” into sequences of length n , and each such sequence α_i is transmitted through the channel $[A_0, \lambda_0, B]$, giving at the output a sequence of $n+m$ letters from the alphabet A . The last n letters of this sequence form the “received” sequences β_k . The output of our transmission process consists of a sequence of such sequences. Our problem is to estimate the rate of transmission. To do this, we consider a sequence of length $s=nt+r$ from the output of the source $[A_0, \mu]$, where t is any positive integer and $0 \leq r < n$. We denote such a sequence by X and the set (space) of all such sequences (for a given s) by $\{X\}$. Let the s -term sequence Y (of letters from the alphabet B) be received at the channel output, when the sequence X is transmitted through the channel $[A_0, \lambda_0, B]$, and let $\{Y\}$ denote the set (space) of all such sequences Y . We denote by $H_Y(X)$ the conditional entropy of the space $\{X\}$, averaged over Y .

Each sequence X of length $s=nt+r$ can be decomposed into t consecutive sequences $\alpha^{(1)}, \alpha^{(2)}, \dots, \alpha^{(t)}$ of length n and a “residual” sequence α^* of length $r < n$. Clearly we can regard

the space $\{X\}$ as the product of the $t+1$ spaces $\{\alpha^{(j)}\}$ ($1 \leq j \leq t$) and $\{\alpha^*\}$, where each of the first t spaces has the structure of the space which we considered above. It is obvious that in general the $t+1$ spaces will be mutually dependent. Because of the basic property of the entropy of a product space, we have

$$H_{Y_0}(X) \leq \sum_{j=1}^t H_{Y_0}[\alpha^{(j)}] + H_{Y_0}(\alpha^*),$$

where Y_0 is any fixed sequence of Y . Thus, averaging over Y_0 , we find

$$(16.7) \quad H_Y(X) \leq \sum_{j=1}^t H_Y[\alpha^{(j)}] + H_Y(\alpha^*).$$

Just as we decomposed the sequence X into t sequences $\alpha^{(j)}$ and the residual sequence α^* , we can decompose the sequence Y into t sequences $\beta^{(j)}$ ($1 \leq j \leq t$) of length n and a residual sequence β^* of length $r < n$. Then the space $\{Y\}$ will be the product of the $t+1$ spaces $\{\beta^{(j)}\}$ ($1 \leq j \leq t$) and $\{\beta^*\}$. The sequence $\beta^{(j)}$ ($1 \leq j \leq t$) is the sequence at the channel output corresponding to the sequence $\alpha^{(j)}$ at its input. Therefore the product space $\{\alpha^{(j)}, \beta^{(j)}\}$ ($1 \leq j \leq t$) has the distribution $\omega(\alpha_i, \beta_k)$ considered above, and the space $\{\beta_k^{(j)}\}$ has the distribution $\eta(\beta_k)$. Let $B^{(j)}$ ($1 \leq j \leq t$) denote the set of all the sequences $\beta^{(l)}$ ($1 \leq l \leq t$) and β^* which make up Y , *with the exception of* $\beta^{(j)}$. Then in order to fix a specific sequence Y , we must fix $\beta^{(j)}$ and $B^{(j)}$. In other words, we can regard the space $\{Y\}$ as the product of the spaces $\{\beta^{(j)}\}$ and $B^{(j)}$. (Here j is any of the numbers $1, 2, \dots, t$.) Therefore, by Lemma 1.2, we find for any j ($1 \leq j \leq t$)

$$(16.8) \quad H_Y[\alpha^{(j)}] = H_{\beta^{(j)} B^{(j)}}[\alpha^{(j)}] \leq H_{B^{(j)}}[\alpha^{(j)}] = H_B(\alpha).$$

On the other hand, the space $\{\alpha^*\}$ obviously contains a^r events,

where a is the number of letters in the alphabet A_0 . But by one of the fundamental properties of the entropy the entropy of a finite space does not exceed the logarithm of the number of events in the space (see [1], [5]). Therefore, for any choice of the sequence Y_0

$$H_{Y_0}(\alpha^*) \leq r \lg a < n \lg a$$

and, consequently, the averaged conditional entropy $H_Y(\alpha^*)$ satisfies

$$(16.9) \quad H_Y(\alpha^*) < n \lg a.$$

Then by (16.8) and (16.9), the inequality (16.7) gives

$$H_Y(X) \leq t H_p(\alpha) + n \lg a,$$

from which it follows by (16.6), that for arbitrarily small $\lambda > 0$, for sufficiently large n , and for any $t \geq 1$

$$H_Y(X) < \lambda tn + n \lg a \leq \lambda s + n \lg a.$$

Now we recall that in #10 we characterized the quantity $H_Y(X)$, the “residual entropy” of the sequence X after it has been transmitted through the channel, as the amount of information remaining in the sequence after transmission, i.e., lost during transmission. Since the amount of information contained in the sequence X prior to transmission is sH_0 , the amount of information transmitted is $sH_0 - H_Y(X)$. But transmitting each sequence $\alpha^{(j)}$ (of length n) requires the passage of $n+m$ symbols through the channel, and just as many symbols are necessary for transmitting the sequence α^* . Therefore the total number of symbols passing through the channel during the transmission of the sequence X is $(t+1)(n+m)$. Thus, one symbol at the channel output carries on the average an

amount of information

$$\frac{sH_0 - H_r(X)}{(t+1)(n+m)} \geq \frac{sH_0 - \lambda s - n \lg a}{n(t+1)\left(1 + \frac{m}{n}\right)} \geq \frac{sH_0 - \lambda s - n \lg a}{(s+n)\left(1 + \frac{m}{n}\right)} = \frac{H_0 - \lambda - \frac{n \lg a}{s}}{\left(1 + \frac{n}{s}\right)\left(1 + \frac{m}{n}\right)}.$$

If we select n large enough so that $\frac{m}{n} < \varepsilon$ and then t large enough so that $\frac{n}{s} \leq \frac{1}{t} < \varepsilon$, then the right side will be larger than

$$\frac{H_0 - \lambda - \varepsilon \lg a}{(1 + \varepsilon)^2} < H_0 - 2\lambda,$$

if ε sufficiently small. This very important result means that with the coding selected each letter received at the channel output brings on the average an amount of information as close as desired to that which one letter carries on the average at the source output. In other words, the transmission of information occurs at a rate arbitrarily close to that at which the information is emitted by the driving source. Of course, all this is only under the condition that $H_0 < C$, and for coding of sufficiently long sequences. In this regard, we note that if n must be taken too large, then the practical value of the coding method described is nullified, since one would have to wait too long to decode (decipher) the text received at the channel output. Thus, from the practical point of view, it would be of considerable interest to investigate the relation between n and λ . Feinstein [4] obtained some interesting results in this direction, but we shall not go into them here. From the purely practical point of view, we must note again that in both the Feinstein and the Shannon methods the construction of a code with the required characteristics is not given; the existence of

such a code is proved, but no indication is given of how to actually find it.

The result which we have obtained is the second Shannon theorem, and can be stated as follows.

Theorem.

Under the conditions of the theorem of #15, there exists a code such that the rate of transmission is as close to H_0 as desired.

Conclusion

The proof given above, which is based on the ideas of Shannon, McMillan, and Feinstein, is certainly long and complicated. However, it rests entirely on a single central idea which, when correctly understood, makes very transparent all the different steps of the complicated demonstration. Therefore, we consider it expedient to stress again, in somewhat more detail, the basic idea which lies behind all our considerations.

The channel we are given is a noisy channel. This means that we cannot determine the sequences of symbols sent at the channel input from the sequences received at the channel output; because of noise, two different sequences at the channel input can give rise to the same sequence at the output. The situation is considerably improved if we know two groups B_1 and B_2 of n -term sequences at the channel output such that 1) B_1 and B_2 do not contain any sequence in common, and 2) with overwhelming probability, the first of the two sequences which might be sent is mapped into one of the sequences of the group B_1 , and the second into one of the sequences of the group B_2 . In this case, by receiving a sequence of group B_1 at the channel output, we can almost be assured that the message was the

first of the two possible sequences, whereas if a sequence of the group B_2 is received at the output, then it is almost certain that the second sequence was the message. Thus, under the circumstances which we have described, these two transmitted sequences are *distinguishable*. Groups of three, four, and more distinguishable sequences (as yet we do not touch upon the question of the existence of such groups) can be determined at the channel input in exactly the same way. Suppose we were able to find a group consisting of a large number K of such distinguishable n -term sequences at the channel input. If we could limit ourselves to sending only sequences belonging to this group, then from the sequence received at the channel output, we would be able to determine almost without error the sequence transmitted.

Can we do this? Obviously, in order to do so, it is necessary that the number L of different n -term sequences from the output of the given source which it is required to transmit through the channel, should not exceed K ; for under this condition (and only under this condition) will it be possible for us to code the whole group of L such sequences which might have to be transmitted into our "distinguishable" group of K sequences at the channel input. Thus, the inequality $L < K$ serves as a criterion of the possibility of transmitting almost without error, and our efforts must be directed towards making L as small as possible and K as large as possible. The first goal is attained by using McMillan's theorem (Ch. II). By neglecting the "low probability" group (with total probability as small as desired) and, consequently, restricting ourselves to transmitting sequences from the "high probability" group, we greatly reduce the number of these sequences, and make it

approximately equal to 2^{nH_0} , where H_0 is the entropy of the given source. The second problem is solved using Feinstein's fundamental lemma (Ch. IV), which asserts the existence of distinguishable groups for which the number of terms is $K > 2^{n(C-\varepsilon)}$, where C is the (ergodic) capacity of the channel, and ε is a positive number as small as desired. If $H_0 < C$, then we see at once that we have $L < K$ for a suitable choice of the distinguishable groups, and our problem is solved.

This is the central idea of the proof. All the rest is merely technique, although this technique sometimes requires great ingenuity in overcoming difficulties which arise.

In all the formulations of the Shannon theorems which exist in the literature, these theorems are accompanied by the converse proposition: if $H_0 > C$, then coding with the required effect is impossible. All authors regard this statement as almost obvious and allot its proof only several lines. In the version of the proof which is given in this paper, I do not see the possibility of proving these converse propositions. The whole matter here is the definition of C , a definition which, in my opinion, is usually given rather carelessly. It is said that the capacity C of a given channel is the least upper bound of the rate of transmission over this channel of the output of all possible sources (the alphabet of which coincides with the input alphabet of the channel.) As far as I can see, here the words "all possible sources" must be replaced by the words "all possible ergodic sources". (Consequently, the capacity which I define is called "ergodic".) Without this addition, the proofs of both McMillan and Feinstein break down at their central point. But if C is understood to be the ergodic capacity of the given channel (as is done in this paper) then the converse propositions

mentioned above are not only not obvious, but apparently require essentially new ideas for their proof. I consider it possible that this difficulty can be overcome by using known limitations on admissible coding systems.

References

1. C. E. Shannon, *The mathematical theory of communication*, Bell Syst. Techn. Journ., **27**, 379-423; 623-656 (1948).
2. S. Goldman, *Information Theory*, Prentice-Hall, New York, (1954).
3. B. McMillan, *The basic theorems of information theory*, Ann. Math. Statistics, **24**, 196-219 (1953).
4. A. Feinstein, *A new basic theorem of information theory*, Trans. I.R.E., PGIT, 2-22, September (1954); identical with Technical Report No. 282, June (1954), Research Lab. of Electronics, Mass. Inst. of Technology.
5. A. Ya. Khinchin, *The entropy concept in probability theory*, the first of the papers translated here.
6. Hardy, Littlewood, and Pólya, *Inequalities*, Cambridge University Press (1934).
7. J. L. Doob, *Stochastic Processes*, John Wiley, New York (1953.)
8. E. Hopf, *Ergodentheorie*, Springer, Berlin (1937); reprinted by Chelsea Publishing Co., New York, 1948.
9. N. Wiener, Duke Math. Journ., **15**, 1-18 (1948).

(Translated by R. A. Silverman and M. D. Friedman)

THE MATHEMATICAL FOUNDATIONS OF INFORMATION THEORY

by A. I. Khinchin

The first comprehensive introduction to information theory, this book places the work begun by Shannon and continued by McMillan, Feinstein and Khinchin on a rigorous mathematical basis. For the first time, mathematicians, statisticians, physicists, cyberneticists and communications engineers are offered a lucid, comprehensive introduction to this rapidly growing field.

Nearly ten years ago, in 1948, Shannon published his original investigations. Since then, a number of specialized papers and one non-rigorous treatment have appeared but this book is the only complete and exact introduction to the subject.

In his first paper, Dr. Khinchin develops the concept of entropy in probability theory as a measure of uncertainty of a finite 'scheme,' and discusses a simple application to coding theory. The second paper investigates the restrictions previously placed on the study of sources, channels and codes and attempts, "to give a complete, detailed proof of both . . . Shannon theorems, assuming any ergodic source and any stationary channel with a finite memory."

PARTIAL CONTENTS: I. **The Entropy Concept in Probability Theory.** Entropy of Finite Schemes. The Uniqueness Theorem. Entropy of Markov chains. Application to Coding Theory. II. **On the Fundamental Theorems of Information Theory.** Two generalizations of Shannon's inequality. Three inequalities of Feinstein. Concept of a source. Stationarity. Entropy. Ergodic sources. The E property. The martingale concept. Noise. Anticipation and memory. Connection of the channel to the source. Feinstein's Fundamental Lemma. Coding. The first Shannon theorem. The second Shannon theorem.

Translated by R. A. Silverman and M. D. Friedman. Bibliography. vii + 120pp. 5 3/8 x 8.

TF1 Paperbound \$1.35